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## About an Argument of Newton


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The view is first developed that it was in the effort of Copernicus' successors to justify fundamentally the validity of the change of conception usually associated with his name that the science of dynamics took a systematic shape. Then Newton's argument is examined that the ptolemaic alternative must mean a disruption of the solar system which is not witnessed and, therefore, the Copernicus-Kepler scheme ought to be accepted absolutely. This particular reasoning of Newton is shown, however, to suffer from a defect. The basis of others of Newton's arguments on the subject is considered, and it is presented that the application of dynamical principles makes the copernican standpoint on planetary motions no more final than the high degree of probability adduced in its favor by Copernicus himself on the ground of its merit of simplicity and harmony as compared with Ptolemy's picture.


THE elements of astronomy featured as heliocentric were proposed by Copernicus[1]; I say "elements" of such astronomy rather very intentionally for it turns out that in Copernicus' scheme, for purpose of certain refined computations that I cannot detail in this place (see Book III of Copernicus' *De Revolutionibus*), the center of the orbit of the earth, even while it does not represent the location of the sun, could be yet accorded a kind of "central" importance.

The shift from the viewpoint called ptolemaic to that named copernican was beset with profound difficulties. It was, in the opinion of the writer, in the search for rational and radical justifications for the insight or belief in the essential correctness in this shift of viewpoint that the whole science of motion, comprised in kinematics and dynamics, and of allied subjects attained to a coherent maturity in the celebrated epoch from 1543 to 1687, from the year of the publication of Copernicus' *De Revolutionibus* to the appearance of Newton's *Principia.* And a notably striking appreciation of this development is obtainable, I think, if we go carefully into the ideas in the sequence more or less in which they emerged. Accordingly, I shall commence in that order. But in so far as the culmination of these trains of reasonings was reached in an effort of Newton (and also because I reconsider Newton's argument), I have chosen the heading of "About an argument of Newton" for this piece of writing.

Now the ptolemaic viewpoint had the incalculable authority of an appeal to direct appearances, and it also enabled such appearances to be generally "saved" (I do not here regard the position that *σώζειν τά φαινόμενα* (só zeintá phainomena) may not be too literally put down as "saving the appearances"). To entertain the notion of abandoning it and of accepting another viewpoint which has no such direct support, rather the contrary, at least two basic conditions

[1] I have no use here for the earlier conjectures or formulations of his remote or immediate predecessors.

must at once be fulfilled. These conditions are, first, that in the realm of motions a nonobvious view need not be ruled out directly and, second, that there are weighty reasons for inclining in its favor. One could, then, further go on to examine circumstances, in a kinematical, dynamical, or indeed in any other sense, the existence or nonexistence of which could provide a really decisive argument. This would clinch the whole matter.

Copernicus complies with these two basic conditions in this way. He first points out the reciprocity of apparent motions (I,5 of *De Revolutionibus*). This permits us, to begin with, to consider the heliocentric viewpoint. Now the germ of the whole problem is already detectable in this very notion of Copernicus. If one change of position is to be equivalent to another (corresponding) change of position so that the same consequence is manifest in either case, how are we to label the one or the other as "real" or merely "apparent" and so decide the problem? This brings us to Copernicus' second line of argument. He applies this reciprocity of manifest motions to planetary phenomena (see I,9), clears away many of the traditional objections to the mobility of the earth, and exhibits the heliocentric standpoint, while the appearances are being effectively "saved," as far less complex in every way. The emphasis is on the simplicity of the heliocentric picture. But Copernicus fully recognizes the obstacle inherent in an attempt at decision—any merely "apparent" change of position is equal to a corresponding "real" change of position after all—and summing up in Book I of his *De Revolutionibus* he concludes that earth's motion (rather than the sun's) is more probable. The merit of simplicity thus only creates a marked probability in favor of the copernican standpoint. A clearly decisive verdict is not obtainable yet.

The inherent difficulties of a decision in the matter on kinematical lines find an expression also in the justified ease with which the renowned Tycho Brahe, that brave Danish nobleman with whom modern observational astronomy began its earnest rise, proposes his well-known amendments to Copernicus' design. We come next to the great name of Kepler. His work augments directly or indirectly the content of this "simplicity." He further suggests to Galilei to enlist the aid of the direct trigonometric methods of determining the parallaxes, then seldom practicable in view of the excessive largeness of astronomical distances relative to the observational technique of the time, for arriving at the parallactic motion expected on copernican theory and so affording it a very strong support. It is clear to him that the distances of the stars as determined from the intensities of the light would be, of course, practically the same whichever planetary system were the true one and would therefore be of no help for furtherance of the objective. This suggestion to Galilei was a notable proposal of a kinematical nature.[2] But he perceives with a truly incomparable insight, for the proper "laws" of motion were not discerned or differentiated by then, that kinematical reasonings alone would be inadequate for the purpose and that the true grounds for arbitration were rather to be sought in causal physical factors, "the inner astronomy," as he beautifully calls it (*Proem. Astronomiae Pars Optica*, 1640, Op. t.ii). From Kepler's time onwards we encounter an acute search into the nature of this "inner astronomy" and the related problem of motion in general, to understand the relation of physical circumstance to the varied manifestations of motion. It is this turn in scientific thought which, to my mind, organically binds together the otherwise apparently, from a contemporary point of view, disjointed work of Kepler and Galilei. Galilei applies his knowledge of the nature of inertia of matter, extracted with unexampled intuition, to an explanation of the tides in such a manner as to exhibit them as a direct result of the truth of copernican plan. The complete explanation escapes his penetrating sagacity, it depended on a huge superstructure while the foundation for that he was yet laying, though he rightly deems the tides to arise from relative differences in motions concerned. He also considers, on the kinematic side, his discoveries of Jupiter's satellites and the

[2] How pressing seems to have been the feelings of the times for a decision in this matter, and how difficult at the same time it was, are evidence that he is compelled even to make use of the then universally admitted existence of planetary souls as one argument against the ptolemaic viewpoint. How distracted, he says, must be the poor spirits who "*ad tam multa respicere jubenter ut planetam duobus permixtis motibus invehant!*" *Opera*, ed. Frisch, vol. iii, p. 144.

phases of Venus as of great weight in the decision between the system of Copernicus and that of Ptolemy. But being the first of avowed experimental philosophers that he was, he seemingly prefers to delay general application of his dynamical information, abstracted from accessible experience, to every variety of motions and their problems that may obtain in the heavens or on the earth (even if the formal mathematical apparatus was equal to the tasks).

To venture on such a profound piece of daring, and also act of faith, France gave birth to her ablest son, aptly designated by that rare spirit Emile Meyerson, in his *Identity and Reality*, as "the most powerful mind that humanity can boast of," the celebrated René Descartes. He reorganized the foundations of all European thought (and thereby determined the frame of the world's modern civilization); and also brought about the correlation of epistemological questionings, mathematical deduction, and content of phenomena, which has remained the heuristic guidance of scientific inquiry ever since his time. And his much misunderstood theory of vortices, the structure of which cannot be looked into in this place, provided the first attempt based on comprehensive and universalized physical reasoning for labeling the changes of planetary positions judged from the copernican standpoint as the "real" ones.[3] I must, however, say a few words on what seem to me to be the true, but unluckily hardly recognized, subtlety and fertility of Descartes' theory. The essential feature that distinguishes, in my opinion, Descartes' theory from the doctrine of central forces is what I mention hereafter, and not, so I believe, the comparatively superficial question of a "transmission of impulse by a vortical contact" or by "gravitational action at a distance." The true fundamental point, it seems to me, is that Descartes assigns primary importance to happenings taking place in the regions between the bodies, and bodies are, so to say, merely a by-product of the intervening vortices—mere "singularities," as we might say today, in a corresponding context. The doctrine of central forces, as the name implies, gives, on the other hand, a central significance to the bodies and locates the primary seat of happenings in them. Fundamental developments in physical conceptions since the time particularly of Clerk Maxwell are returning to this essential feature in Descartes' line of speculation (to the extent of the underlying notion as I have put it above).

We now approach Newton, whose efforts represent a culmination of the attempts to decide conclusively between the copernican standpoint and that which goes by the name of ptolemaic. In order to deal at some length with this culminating effort, I find it useful to recapitulate in a few words and then to gain admittance to Newton's arguments.

The heliocentric astronomy was proposed by Copernicus on the grounds of the relativity of motion and of the higher probability of the existence of an arrangement possessing a simpler order. Copernicus' successors strengthened his reasonings by highly important simplifications of a kinematical nature, and it was also foreseen from the divinations of Kepler onwards that a basis for decisive arbitration would have rather to be sought in the physical factors that actuate or influence the motions. But a consistent scheme of mechanics was not attained and the inferences from such mechanics as were surmised rather reflected the embryonic stage through which that subject was passing.

When Newton took up the task of elaborating the mechanics of the heavenly bodies in the "solar system," he proceeded to examine if his systematized dynamical methods could furnish a definite verdict. Newton's reasoning appears to branch off into two parts. On the one hand, he observes that the facts pertaining to planetary movements call for an application of his systematized dynamical principles, and he exhibits the extraordinary appropriateness of the Copernicus-Kepler picture for an application of his scheme of mechanics (an application set out in detail in the various *Phaenomena* and subsequent Propositions referring to them in Book III of his *Principia*). On the other hand, he shows that in the light of these principles and their application, the opposite conception would entail a profound disaster to the planetary system, which we never witness. The quotations from his works that I reproduce here seem to show that it was very much the nonexistence of the

[3] The sophistry introduced by him in interpretations, to avoid the Church's censure, is of no account for our purpose.

latter contingency that would appeal to him as a decisive argument.

The development of Newton's argument is, of course, best given, here in its substance only, in his own words; and to this end paragraphs abstracted from pp. 1–28 of Motte's translation of Newton's *System of the World* are included here.

> Whence it was that the planets came to be retained within any certain bounds in these free spaces, and to be drawn off from rectilinear courses, which left to themselves, they should have pursued, into regular revolutions in curvilinear orbits, are questions which we do not know how the ancients explained; ···for from the laws of motion it is most certain that these effects proceed from the action of some force or another···but by the action of a force (circum-solar) so great it is unavoidable that all bodies within the bounds of the planetary system must descend directly to the sun, unless by other motions they are impelled towards other parts···since no such thing happens it must needs be they are moved towards other parts; and nor is every motion sufficient for this purpose. To hinder such a descent a due proportion of velocity is required···and therefore the planets, Saturn, Jupiter, Mars, Venus and Mercury, are not really retarded in their perigees nor become really stationary, or regressive with slow motions. All these are but apparent, and the absolute motions, by which the planets continue to revolve in their orbits, are always direct, and nearly equable. But that such motions are performed about the sun, we have already proved; and therefore the sun, as the centre of the absolute motions, is quiescent. For we can *by no means* allow quiescence to the earth lest the planets in their perigees should indeed be truly retarded, and become truly stationary and regressive and so for want of motion should descend to the sun···.

With regard to this drastic argument of Newton, I fear there was a sort of oversight here; it appears there was a confusion between the thought of the sun and these planetary bodies confronting one another in a condition of common rest (and nearly so) for periods long enough to be critical and the quiescence of the earth as opposed to the quiescence of the sun. When we accord true quiescence to the earth though these planets indeed then manifest true retardations in their perigees, become at points really stationary (and regressive), yet they cannot for all that descend to the sun. Because the sun thereby comes to acquire a correspondingly true motion which saves them from direct descent through effective compensations by the consequent displacements of the sun. A little further reflection will show that although the sun only acquires a fixed velocity (and fixed changes of velocity), the situation of the "due proportion" of true velocity which a planet possessed on the first view and in virtue of which it was prevented from dropping down to the body of the sun finds an adequate counterpart in each individual case through the agency of movements of the sun.[4] I say to "the body of the sun" for precision; we must explicitly refer in this connection to the sizes of the planetary bodies. For as regards a point-center of force operating under the inverse square law a "point-particle" will by any velocity be able always to describe some conic about it, or respecting it, in free spaces, unless it happens to move along the right line connecting it and the point-center (Prop. XIII, Cor. 1, and Prop. XVII, Book I, *Principia*). However, I must at the same time add that although the sun and planets are not mere points but are large bodies, still their distances apart are so enormous compared with their radii that we can treat them simply as massive points and the mention of "the body of the sun" is, as indicated with such mention, for purpose of precision only. To resume, and using a rather vivid and even light language, it seems that when the sun lies in wait for the planets they are fortunately or unfortunately far too quick to be able to seize opportunity to meet him and therefore overshoot the mark, and that when they somehow "overcome" this drawback, the sun starts correspondingly to play the truant effectively and thus avoids their descent to and union with him.

The size and shape of the earth's orbit and its range of distance from the sun are such that over a small fraction of the length of its orbit the centripetal acceleration can be regarded without sensible error as an acceleration along parallel directions and of unvarying quantity. Therefore, for a small fraction of the length of the orbit, we can speak of earth's motion as compounded of uniform rectilinear motion (along a corresponding particular tangential direction) and of constant acceleration in corresponding (very nearly) parallel directions.[5] And the part

[4] Fuller treatment follows in the next paragraph.

[5] I.e., instead of the actual slightly elliptic shape of such portion of the orbit we speak of it as a slightly rather parabolic one, the two of which can be considered, within the range indicated, as more or less the same. It is to be

of it being traversed during the interval when the five planets are in their perigees is such a small fraction. Let us impress on the planetary system a speed equal to but oppositely directed to the earth's velocity,[6] its tangential velocity at a moment when the other planets are considered to be entering the ("dangerous") zone of the perigees, and also an acceleration in value equal to but directed in direction contrary to the corresponding (very nearly) parallel directions in which it can be considered as experiencing this to the extremely high degree of approximation. The earth, then, will be reduced to rest during this short interval. (***And the assumption of the real quiescence of the earth means nothing more or less than this position outlined above, to the high degree of approximation.***) But in virtue of Cors. V and VI to the Axioms or Laws of Motion (Book I of Newton's ***Principia***) all the motions in the planetary system will happen after the same manner as when the earth was having its heliocentric motions. And as there was to be no question of direct descent to the sun in that case, none can now emerge in the case of the quiescence of the earth. When, of course, the planets recede from the perigees, and are away from them, their velocities relative to the shifting sun, as estimated from a quiescent earth, will be so varied that a sort of "solar system" arrangement (of more or less Tycho Brahe's pattern), though one moving and varying in shape, could in a measure be accorded to them.

One might be tempted now to raise the following conjecture: The assumption of the earth's quiescence and conditions allowing an application of Cors. V and VI are indeed almost equivalent positions, within the sphere indicated; and, therefore, Newton's argument over the crash of the five primary planets into the sun is of course entirely unacceptable. But yet the two positions are not totally equivalent, however slight the difference may be. This difference, be it however small, would provoke variations in dynamic effects, again however slight, as compared to the heliocentric position. We might therefore, instead of watching for the doom of the planets, be on the lookout rather for certain minute changes in the elements of the orbits of the five primary planets, the absence or existence of which could be used to arbitrate between Copernicus and Ptolemy (of course this reasoning has meaning only as a refinement of Newton's particular argument).

I do not, however, go into this question here; I merely observe that it is incorrect to argue at all in this fashion. The whole of these lines of arguments rest on the idea, dating to the time of Copernicus when, owing to the absence of the expected parallax, he rightly placed the fixed stars at their enormous remoteness from our regions, that we can, in effect, abstract the motions of planetary bodies from the rest of the universe and deal with them as if they formed a completely individual system. This trait in the kinematical picture found its way into the dynamics with the resultant ever nicer efforts, much after the fashion of medieval schoolmen-logicians, to concentrate on the narrow question of deciding between the "rest" of the sun or "rest" of the earth, but this time by invocation to forces.

I need hardly hasten to add that in the above discussion I concern myself primarily with Newton's specific argument advanced within his system that as the criterion of the earth's quiescence involves a dissolution of the "solar system" arrangements which we never witness, it must for this reason be ruled out altogether. The problem how we are to furnish also satisfactory dynamical grounds for the alterations in the sun's motions, or for that matter for the general features of the planetary movements (and indeed of wider motions), that the earth's quiescence would require is strictly not germane to a particularized discussion. I take it up below as a fragment in a generalized discussion, though I begin from it.

But any complexities encountered on this account serve principally to emphasize what I called the first branch of Newton's reasoning,

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observed that I do not here speak of the parabolic shape in the sense of Prop. XVII, Book I, which depends on the velocity, but in the sense of (almost) uniform accelerative gravitation acting along (almost) parallel directions for a small part of the tract situated at an immense distance from the center of force. The consideration is similar to that which allows us to regard the trajectories of projectiles on the earth as (almost) parabolic, i.e., the sense is that in which Newton derives what he calls "Galileo's theorem" about the parabolic path of terrestrial projectiles (in a Scholium in his Book I).

[6] Of course we are here dealing out from a heliocentric attitude.

namely, the extraordinary appropriateness of the Copernicus-Kepler picture for his dynamical ideas. I shall therefore deal mainly with this point only. That this would not take the matter beyond what may be termed the "copernican scale of probability" is, I think, rightly manifest from the connected general considerations into which the subject merges, which considerations can only be touched upon here.

Any special importance which Newton's dynamical views themselves and their application to the actual universe may have arises solely from the fact that they constitute, too, a simple description. To describe the motions of bodies which form our solar system, we may either, as in the ptolemaic system, specify the motion of these bodies relative to the earth, or, as in the copernican system, specify the motion of these bodies relative to the sun. Either description (provided its details are correct) is equally valid and the principal ground in favor of the latter is of course, as Copernicus himself argued, its simpler orderliness. Newton only discovered and applied a yet simpler manner of description, by inventing a comprehensive theory of dynamics which follows logically from three fundamental "laws." However, his theory and theorems based on it are still descriptions, in terms of postulate of forces, their dependence and proportions, and their effects. They do not explain anything for the actual world as a fundamental necessity, a conformity with which (e.g., the equal description of areas in equal times about a center of force in free spaces is answered by the Copernicus-Kepler picture) would in itself represent an inherently true position and a variance with which (e.g., the theorem as to the description of areas is not echoed in the ptolemaic design) would correspond to an unacceptable position. For whilst his "laws" of motion are logically incapable of proof as inductions from experimental facts without involving in the process these very laws themselves, they cannot yet by any stretch of imagination be treated as characterized by or having the meaning or import of self-evidency. They partake, therefore, essentially of the character of free specifications, useful for compact description of phenomena of motion, but such specifications indeed that very numerous deductions therefrom accord with a very wide range of observed phenomena. And further, the application of this in itself simple scheme of description of motion to the copernican standpoint is, at the same time, far more straightforward and less complex than to the ptolemaic.

I find it somewhat pointless to debate the question in which sense Newton would have taken the comments along these lines of the famous physicist-philosophers Kirchhoff and Mach. A refinement of these comments, however, points essentially to the position that forces need have no intrinsic intensities or directions. Now Mach hints,[7] though in a somewhat different context, that when Newton said "*hypotheses non fingo*" he may have meant, among other things, that such terms in vocabulary of mechanics as centripetal force, centrifugal force, etc., he did not employ in as serious a sense as we have come to ascribe to them (through conferring on them that immediacy of evidence which pertains to certain muscular sensations).

However that be, the point we draw upon here is that the newtonian shape of the principles of mechanics, that is to say, these assumptions or "laws" and mathematical formulations based on them, and their application to the actual universe, in particular to the Copernicus-Kepler specification of planetary motions, are exactly like copernican astronomy, only simpler for very many of the factors presented. We may regard these principles and their relevant application as confirming Copernicus' standpoint, ***but within his "scale of probability."***

It is, I believe, almost a duty of anyone alluding to epistemological aspects of scientific thought to acknowledge the debt one owes to the luminous writings of the philosopher Emile Meyerson. In my case it is indeed a most charming obligation to do so, for nothing perhaps reveals in an equal degree the unique nobility of his nature than the fostering interest he took in the progress of my humble studies, in spite of age and infirmity, and from across the continents.

[7] *Science of Mechanics* (Open Court, 2nd Eng. edition, 1902), p. 183.

# High Energy Nucleon-Nucleon Scattering Experiments at Berkeley

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The experimental results for $n-p$ and $p-p$ scattering at high energies are described, and a review of the present status of theoretical interpretation is given. For the scattering of 90-Mev neutrons by protons the angular distribution indicates that only even states of angular momentum are strongly effective. Evidence of the operation of exchange forces is mentioned. In the scattering of 32-Mev protons by protons the angular distribution is appropriate to $s$-state interaction alone. The 340-Mev $p-p$ results indicate the existence of noncentral (tensor) forces in states of odd orbital angular momentum.

ONE of the ways to study the forces which hold a nucleus together is by means of experiments in which two free nucleons (neutrons or protons) are caused to collide. Nuclei are composed of neutrons and protons bound to each other, and the forces acting between them in the bound state will produce a scattering when they collide as free particles. This principle was first used by Rutherford to investigate with alpha-particle projectiles the outer regions of the force field of a nucleus, which turned out to be purely coulomb and repulsive. Rutherford's success caused nuclear physicists to hope that the fundamental laws underlying the short range characteristically nuclear forces might also be revealed by scattering experiments. This hope tended to be borne out by early work which showed that nuclear forces are attractive and confined to distances of the order of $10^{-13}$ cm. Just because of this short range, however, it was impossible until a few years ago to get more detailed information, because there existed no accelerators capable of producing nucleons with energies greater than about 15 Mev. A nucleon of this energy or less has a wavelength longer than the over-all extension of the nuclear force, so that it cannot be used to study details of the latter. The construction of new accelerators at Berkeley[1] in the University of California Radiation Laboratory, however, made available neutron energies up to 280 Mev and proton energies up to 350 Mev, with correspondingly shorter wavelengths. For this reason, neutron-proton and proton-proton scattering experiments were high on the priority list of the Berkeley research program.

When the Berkeley experiments were begun, there was hope that many features of simplicity would be found, tying together the known properties of nuclei with the conjecture that mesons are responsible for nuclear forces.[2] The results so far obtained have not weakened the position of the meson hypothesis; indeed, they have given it qualitative confirmation. However, the results are certainly not "simple," and it now seems doubtful that they can be used, by themselves, to deduce the complete picture. They remain, of course, as facts which any future theory, meson or other, must explain. In this paper the various high energy scattering experiments will be described, and the interpretation given by the Berkeley theoretical group discussed.

## I. NEUTRON-PROTON SCATTERING

It is not really permissible to use nonrelativistic mechanics to describe neutron-proton scattering at energies of the order of 100 Mev. The velocity of a 100-Mev neutron is nearly half that of light, so that effects of retardation and of mass increase are not negligible. It is expected, however, that great *qualitative* errors will not arise from using the nonrelativistic approximation. For descriptive purposes here, this will suffice although in the actual analysis of the experiments an attempt was made to include at least some of the relativistic effects.

The energy-momentum conditions for a non-

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[1] See the first article in this series, G. F. Chew and B. J. Moyer, Am. J. Phys. **18**, 125 (1950).

[2] For an elementary discussion of the meson theory of nuclear forces, reference can be made to D. L. Falkoff, Am. J. Phys. **18**, 30–39 (1950). It will be assumed that the reader is familiar with the ideas discussed by Dr. Falkoff.

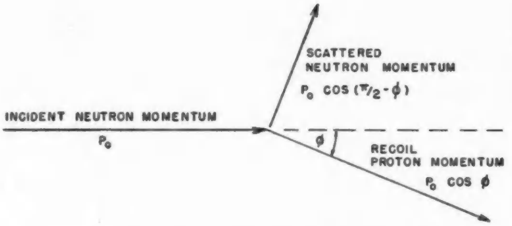


FIG. 1. The momentum relations in a nonrelativistic neutron-proton collision.

relativistic neutron-proton collision are simple because of the near equality of the two masses and can be summarized quite concisely. If the proton is initially at rest and the neutron has a momentum $P_0$, then the final trajectories of neutron and proton are at right angles and the momentum of either is $P_0 \cos\phi$, where $\phi$ is the angle of emission of that particle. (See Fig. 1.) For definiteness we shall henceforth use the symbol $\phi$ to refer to the final *proton* angle, since that is the one observed in the Berkeley experiments. The symbol $\theta$ will be used to denote the final neutron angle in the center-of-mass coordinate system (the system in which neutron and proton approach each other with equal and opposite velocities). It is not hard to show that $\theta=\pi-2\phi$.

The reason protons are bombarded with neutrons instead of vice versa is, of course, that a target of free neutrons cannot be made. The neutron beams which can be obtained from the 184-inch cyclotron are not monoenergetic, as discussed in the previous article on high energy reactions,[3] but this need not be a serious difficulty, since the final momentum of the proton is a measure of the incident neutron energy. A satisfactory experiment, however, requires a knowledge of both the energy and the angle of the recoil protons.

### A. Proportional Counter Measurements with 90- and 40-Mev Neutrons[4]

One method of measuring the energy and angle of the recoil protons is by means of a proportional counter telescope, with supplementary absorbers. The general arrangement of the apparatus which was used for this measurement is shown in Fig. 2. The neutron beam entered from the left and knocked out protons from the scattering target. A telescope of three proportional counters pointing at the scatterer served to define the angle of the recoil protons. Interposed between the last two counters was a copper absorber (A in Fig. 2) of a thickness in the case of 90-Mev neutrons such that it would stop protons of an energy less than 66 $\cos^2\phi$ Mev. This means that all neutrons of an energy less than 66 Mev in the incident beam were ignored when only triple coincidences were recorded. Some protons were lost, also, from scattering in the absorber; but this effect could be corrected for.

To obtain a sufficient concentration of protons

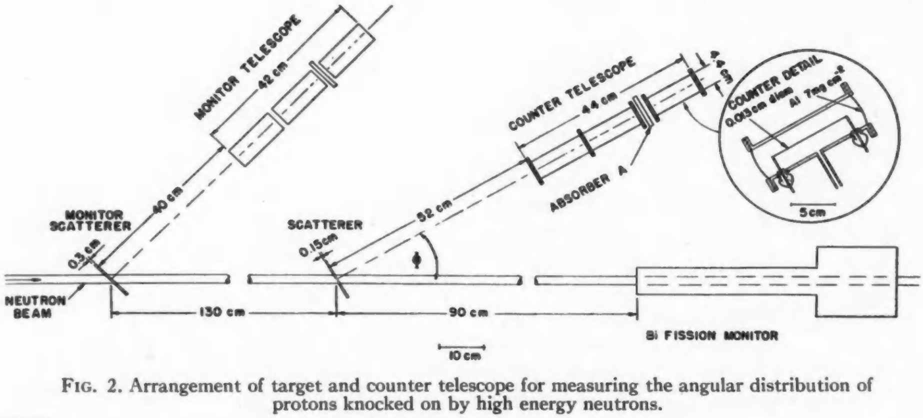


FIG. 2. Arrangement of target and counter telescope for measuring the angular distribution of protons knocked on by high energy neutrons.

[3] B. J. Moyer and G. F. Chew, Am. J. Phys. 19, 17 (1951).
[4] This section is abstracted from a report by Hadley, Kelly, Leith, Segrè, Wiegand, and York, Phys. Rev. 75, 351 (1949).

in the scatterer, it was necessary to use an hydrogenous compound rather than pure hydrogen. The choice made was polyethylene, which has the chemical composition $(CH_2)_n$. By alternating this target with another of pure carbon having the same stopping power and taking differences, it was possible to deduce the scattering due to the hydrogen alone. This is not quite as satisfactory as a direct measurement, and the experiment may eventually be done with a liquid hydrogen target.

The information given by the experiment described was the *relative* angular distribution for the scattering of neutrons whose energy was at least 66 Mev and on the average was 90 Mev. The results are summarized by the lower curve in Fig. 3. The abscissa of this diagram is $\theta=\pi-2\phi$, the final *neutron* angle in the center-of-mass system. (At energies below 15 Mev the angular distribution in this coordinate system is flat as discussed below.) The upper curve of Fig. 3 is the scattered distribution for neutrons of average energy 40 Mev (produced by stripping of 80-Mev deuterons). The various symbols refer to different runs, while the stars give the final best available average. Neither curve extends to 0°, because the proton recoils for such small angles have too small an energy to be detected.

The absolute normalization of the curves in Fig. 3 had to be obtained from an independent determination of the total cross section. This was carried out by a conventional attenuation measurement of the neutron beam in good geometry. This simply means that a neutron flux was measured before and after passing through a hydrocarbon absorber and also a graphite absorber containing the same amount of carbon. The detector subtended a sufficiently small angle so that a negligible number of scattered neutrons were included. From the differences of the three measurements the total cross section of the hydrogen could be deduced. The average values of the total cross section, used in the normalization of Fig. 3 were $0.076\times10^{-24}$ cm² for 90 Mev and $0.170\times10^{-24}$ cm² for 40 Mev. An extrapolation of the curves to zero angle had to be guessed, but because of the small solid angle of the unknown region this did not introduce much uncertainty. More recent unpublished measurements by Leith and Hildebrand indicate that

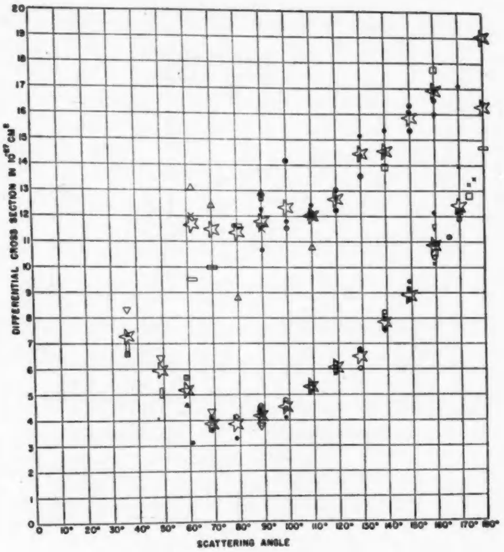


FIG. 3. Angular distribution in the center-of-mass system for neutrons of energy 40 and 90 Mev scattered by protons.

the 40-Mev total cross section may be as high as $0.201\times10^{-24}$ cm².

### B. Cloud-Chamber Measurements with 90-Mev Neutrons[5]

To make sure that no large systematic errors were being made in the counter experiments, a completely independent determination of the 90-Mev $n-p$ angular distribution was carried out with a hydrogen-filled Wilson cloud chamber, placed in a strong magnetic field. The direction and curvature of the proton tracks observed after passing a burst of neutrons through the chamber gave all the required information.

The cloud chamber (Fig. 4) was 16 inches in diameter, with a useful illuminated depth of $3\frac{1}{2}$ inches, and the neutrons entered through a 5-mil aluminum window in the wall of the cylinder. The chamber was filled with hydrogen at a pressure of 110 cm of Hg and saturated with an alcohol-water mixture. A pair of Helmholtz coils supplied the magnetic field of 14,000 gauss, which was pulsed on for 0.15 second every two minutes. During this interval the cloud chamber

[5] Abstracted from Brueckner, Hartsough, Hayward, and Powell, Phys. Rev. **75**, 555 (1949).

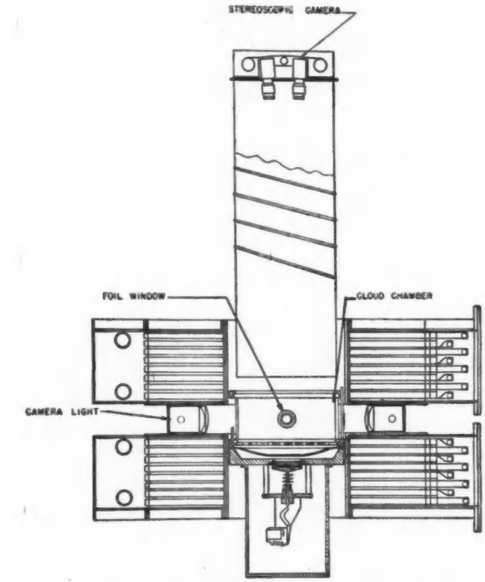


FIG. 4. Diagram of cloud chamber used for 90-Mev $n-p$ scattering experiment.

was expanded and the cyclotron pulsed near the end of the sensitive time of the chamber. Great care was taken to minimize turbulence, the traditional bugaboo of cloud-chamber measurements, and extremely sharp tracks were produced. Figure 5 gives an example. Photographs were taken with a stereocamera and reprojected on a translucent screen. The curvature and direction of all proton tracks, making an angle of less than 84° with the incident neutron beam, were accurately measured. This limit was chosen because at larger angles the tracks are so short that they may sometimes be overlooked.

The momentum of each proton could be deduced from its radius of curvature in the known magnetic field; and, consequently, the energy of the neutron producing the recoil was known (Fig. 1). (This gave an independent determination of the stripping neutron spectrum, which checked well with previous data.) A total of 1764 knock-on protons were counted, excluding those produced by neutrons of less than 40 Mev. Since there were relatively few neutrons between 40 and 66 Mev, this constituted a measurement of essentially the same quantity as in the counter experiment. The data are compiled in the histogram of Fig. 6.

It is seen that the statistical accuracy is not as high as in the measurement with proportional counters, but that the over-all agreement with the latter is good. This fact gives one much more confidence in the counter results, since many unfamiliar phenomena were encountered in this first precision measurement of a high energy scattering cross section.

### C. Discussion of $n-p$ Measurements

In the introduction it was pointed out that when the de Broglie wavelength is longer than the range of the nuclear force, one can derive no information about the details of the latter. This fact manifests itself experimentally in an angular distribution which is spherically symmetric in the center-of-mass system. In other words, low energy measurements of the distribution shown in Figs. 3 and 6 have always yielded a flat function, showing that only the $S$ part of the incident wave was being scattered. The anisotropy of the Berkeley results was the first indication of a nuclear interaction in states of nonzero angular momentum.

A detailed phenomenological analysis of the

FIG. 5. Example of cloud-chamber photograph obtained in the measurement of 90-Mev $n-p$ scattering. Only proton tracks originating in the gas are significant.

high energy $n-p$ experiments has been made by R. S. Christian and E. Hart of the Radiation Laboratory.[6] Their most important conclusion was that the interaction in $p$ states (one unit of orbital angular momentum) of the proton-neutron system is almost absent, most of the observed angular dependence being due to $d$ states (two units). This fact is suggested at once by the near symmetry of the angular distribution about 90°, and it is further borne out by the low total cross section. The result was quite unexpected on the basis of existing ideas about the saturation property of nuclear forces[2] and is far from being understood at the present time.

A more naive interpretation of the results, suggested by meson theory, is to say that the strong maximum at 180° shows the existence of an exchange force. In other words, one obvious mechanism by which a neutron could be found with such a high probability in the backward direction is for the incident particles to have exchanged roles in the process of scattering, the proton becoming a neutron and vice versa. This is probably the case, qualitatively; but the existence of a second peak at 0° shows that ordinary forces are not absent. No reasonable meson theory yet put forward can explain quantitatively the entire angular distribution.

A measurement of the $n-p$ cross section at a still higher energy has also been made at Berkeley.[7] "Knock-out" neutrons, produced by the 350-Mev proton beam impinging on a beryllium target,[3] were used. This measurement, for many reasons, is not so precise as the one already described. The results are qualitatively similar, however, the $p$ scattering still being absent. The central valley of the angular distribution is deeper and the forward peak narrower, but this is to be expected from the higher energy.

## II. PROTON-PROTON SCATTERING

The kinematics of proton-proton scattering are the same as for neutron-proton scattering because again one is dealing with two particles of the same mass. Experimentally, however, there is no way to decide whether the proton observed

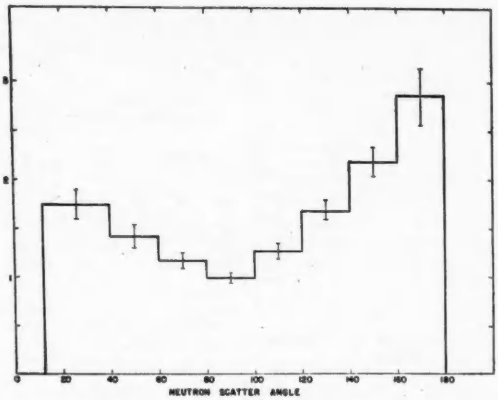


FIG. 6. Histogram compilation of data from cloud chamber experiment.

at a particular angle after the collision is the incident particle or the recoil. The principles of quantum mechanics, indeed, do not allow such a question to be asked and require that the wave function which describes the scattering be antisymmetric in the coordinates of the two protons. The quantity to be measured and to be interpreted is simply the number of protons at a particular angle. The distribution of this number must obviously be symmetric about 90° in the center-of-mass coordinate system, and its integral over all angles is two per collision. The symmetry is always used in a careful experiment as a check on the results.

In addition to this difference between $n-p$ and $p-p$ scattering, which is due to the identity of the two protons, there is the coulomb repulsion which was absent in the $n-p$ case. This produces strong effects at low energies and relatively smaller effects as the energy increases. It is well understood at all energies, however, and can be effectively eliminated in the theoretical analysis, reducing the problem to that of two uncharged protons. Thus, it is a sensible procedure to compare the purely nuclear forces acting in the $n-p$ and $p-p$ systems, respectively, even though the results of the corresponding scattering experiments may look quite different. One of the features of simplicity in the nuclear force picture which had been suggested by low energy results was that the $n-p$ and $p-p$ forces were the same. This, as we shall see, turns out to be far from true.

[6] R. S. Christian and E. Hart, Phys. Rev. 77, 441 (1950).
[7] Kelly, Leith, and Wiegand, Phys. Rev. 75, 589 (1949), and private communication.

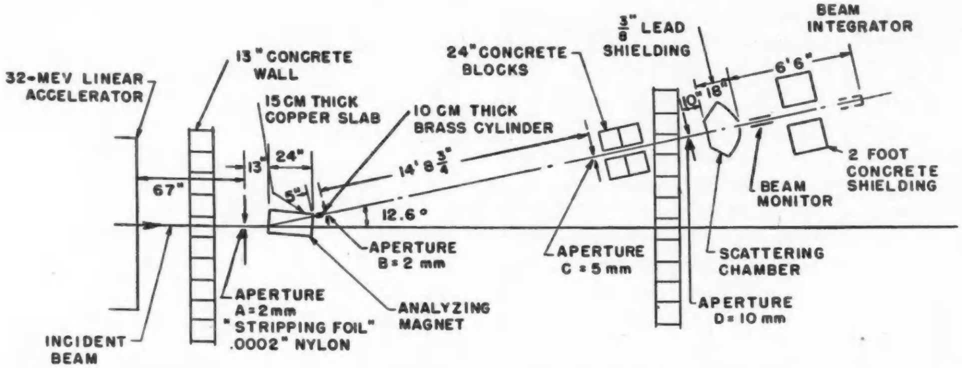


FIG. 7. Scattering chamber and surrounding counters for measurement of 32-Mev proton-proton scattering.

## A. Counter Measurements with Protons from the Linear Accelerator

As described in the first article of this series, the Berkeley linear accelerator is capable of producing an intense, well-collimated, almost monoenergetic beam of protons; and a supplementary steering magnet can be used to define the energy as accurately as is desired. This allowed the measurement of proton-proton scattering near 30 Mev to be more precise than the aforementioned neutron-proton experiments. Because of the importance of this measurement, it was undertaken by two separate groups, using entirely different methods. One group employed counters in an essentially conventional scheme.[8] This method will be described first.

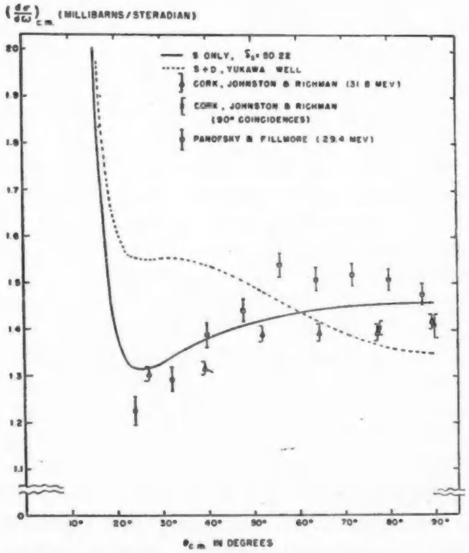


FIG. 8. Results of two independent measurements of $p-p$ scattering near 30 Mev. The slight difference in energy accounts for the different absolute values. The solid and dotted curves are theoretical, showing what is to be expected if nuclear interaction occurs in the $S$ or in both $S$ and $D$ states, respectively.

The experimental setup is outlined in Fig. 7. The proton beam, collimated to a diameter of 7 mm, entered the apparatus from the left through a Nylon window. The entire apparatus was filled with hydrogen gas at atmospheric pressure, although the cylindrical region from which scattering was accepted was only two centimeters long by half a centimeter in diameter. The absolute intensity of the proton current was measured by collecting the charge in a Faraday cup, and seven proportional counters shaped as annular rings recorded the scattering at seven different angles. Only the range 15°–51° in the laboratory system was covered, because of mechanical difficulties in the construction of the apparatus. This meant that the intensity at only one position was checked against the complementary angle intensity.

Background, due to energetic neutrons which produced recoils in the counters, was a problem. These neutrons were generated at a variety of places by the primary proton beam and were greatly reduced by a careful arrangement of absorbers. The remainder, which accounted for less than 10 percent of the counting rate, were

[8] Cork, Johnston, and Richman, Phys. Rev. 79, 71 (1950).

dealt with by the familiar subtraction procedure. This simply meant making alternate runs with and without hydrogen in the apparatus and taking the difference.

The results of the counter measurement are given as the triangles in Fig. 8, plotted in the conventional way against the angle in the center-of-mass coordinate system. Only the range $0-90°$ is shown, because of the symmetry about $90°$. The results will be discussed in Sec. II–D.

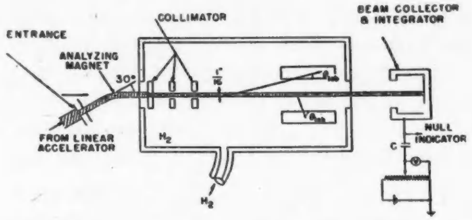


FIG. 9. Diagram of scattering chamber for photographic measurement of $p-p$ scattering at 29.4 Mev.

### B. Photographic Plate Measurements with Protons from the Linear Accelerator

A completely independent measurement of the 30-Mev $p-p$ cross section has also been made with a photographic method of detection.[9] A number of plates were set like fins, edge-on to a small cylindrical volume which was filled with hydrogen. The proton beam from the linear accelerator was allowed to pass through this volume and some of the scattered protons found their way into the emulsion, producing tracks which could be examined under a microscope. The angle which these tracks made with the edge of the plate was the same as the angle of scattering.

Figure 9 shows the chief features of this arrangement. After passage through the magnetic field to improve the definition of the energy, the proton beam was collimated by a succession of three slits, so that it entered the scattering chamber with a width of only $\frac{1}{16}$ of an inch. The near edge of each photographic plate was half an inch from the center of the beam, a picture of the plate holder being shown in Fig. 10.

The chief difficulty with this method arose from slit scattering. Protons may strike the edges of the collimating slits and be deflected enough to get into the emulsion. This circumstance made it impossible to get reliable data at angles less than 10°. Protons observed entering the plates at large angles, of course, must have originated in the hydrogen gas. The maximum observable angle was set by the same consideration as in the $n-p$ experiments: protons at angles greater than 80° have such a small range that their tracks could not be reliably counted.

This experiment, then, gave the angular distribution of protons between 10° and 80° in the laboratory system, a considerable improvement over the range covered by the counter experiment, since the intensity at every angle was checked by the intensity at the complementary angle. Statistically, of course, the counter method had the advantage, being able to record up to five protons per second per counter. A single person is only able to count between 250 and 400 tracks a day in a photographic emulsion; so that 30 man-days are required to accumulate 10,000 counts, the order of magnitude needed for a good measurement of this kind.

With the entire range of angles divided into 8° intervals, an angular distribution was constructed from the 11,000 recorded tracks. Smaller intervals could just as well be chosen, but then the vertical accuracy of each point would suffer. Since the curve is not expected to have sharp

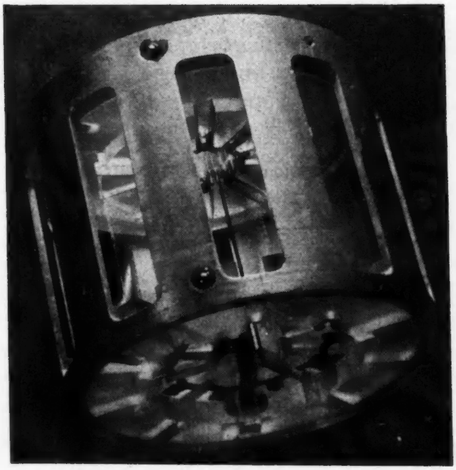

FIG. 10. Photograph of holder for photographic plates used in 29.4-Mev $p-p$ scattering experiment at Berkeley.

[9] W. K. H. Panofsky and F. L. Fillmore, Phys. Rev. 79, 57 (1950).

features at this energy, a rather small number of intervals each with a good statistical accuracy seemed the most sensible choice. The results are shown as the squares in Fig. 8. The agreement between the two independent experiments is seen to be good.

### C. Counter Measurements with 350-Mev Protons from the Synchro-Cyclotron

The proton-proton scattering cross section at 350 Mev has also been measured, using the proton beam from the 184-inch synchro-cyclotron.[10] Once again two different techniques were employed, but both were based on counters. The first technique was almost identical with that described in section I-A for $n-p$ scattering. The same counter telescope, indeed, was used in the two experiments. The second technique again used three counters but in a different arrangement. Two were used without any absorber to form a defining telescope. The third was set at an angle of 90° (actually 85.5° because of relativistic corrections) with respect to this telescope and coincidences between the two positions were required. Coincidences were caused by the two protons emerging from a $p-p$ collision, since they necessarily occur at right angles to each other. Chance simultaneous trippings of the two counters by background radiation were infrequent enough to be ignored. The results of the two measurements were in agreement and extremely surprising. In the center-of-mass coordinate system of the two protons, the scattering cross section was found to be a constant over all the angles measured (40°–90°), with a magnitude of $4.5\times10^{-27}$ cm$^2$ per unit solid angle.

### D. Discussion of $p-p$ Measurements

Almost anything different from the very flat $p-p$ angular distributions observed at 32 and at 350 Mev would have been more understandable. The proton wavelength is sufficiently short already at 32 Mev to make interaction in $p$ and $d$ states possible; but, as Figs. 7 and 8 show, no such interactions manifest themselves in an obvious way. The weak angular dependence observed is easily attributable to the coulomb field. In spite of a similar-looking result at 350 Mev, however, one can be quite sure that scattering does occur there in states of nonzero angular momentum, because the magnitude of the cross section is too great to be due to $s$ scattering only. A rather complicated force scheme for the states of odd orbital angular momentum must be invoked to explain the high flat angular distributions observed at both energies.

The theoretical analysis of the foregoing data, as carried out at Berkeley,[11] cannot be described here; but the conclusions may be stated as follows. (1) The $p-p$ force is not the same as the $n-p$, there being a strong interaction in states of odd orbital angular momentum ($p, f, \cdots$ states) in the former case and only weak if any in the latter. (2) In odd states, the $p-p$ forces have a range much shorter than that observed in even states and do not act along the line joining the particles but are correlated with the spatial orientation of the combined spin of the two protons. These conclusions were reached on the basis of a phenomenological picture of nuclear forces, in which they are derived from non-velocity-dependent potentials. The general conclusions have a good chance of standing up even if the potential model is wrong, but they should not yet be regarded as unequivocal.

That nonspherical terms occur in the nuclear force scheme has been known since the discovery of the quadrupole moment of the deuteron. The $p-p$ results, however, if the interpretation is correct, constitute the first strong manifestation of the asymmetry in scattering phenomena. There is no tie-in with the deuteron, however, because there the effect is in states of even orbital angular momentum, while in proton-proton scattering the effect occurs in odd states.

The lack of similarity between $n-p$ and $p-p$ interactions necessitates no fundamental change in the idea that mesons are responsible for nuclear forces, but it upsets a pleasant feature of simplicity which had almost been accepted among nuclear physicists.

## III. CONCLUSION

It is possible that this is an inappropriate time to discuss the Berkeley experiments, since at the moment the results are so poorly under-

[10] O. Chamberlain and C. Wiegand, Phys. Rev. 79, 81 (1950).

[11] R. S. Christian and H. P. Noyes, Phys. Rev. 79, 85 (1950).

stood. A theory may perhaps be proposed very soon to clarify the situation, but the probability of this seems small. The most widely shared opinion about nuclear forces at present is that they are due to the exchange of mesons between nucleons but that so many mesons are involved, at least in situations so far considered experimentally, that the over-all effect is complicated and cannot be used as the starting point in making a theory. Attention is now turning to the mesons themselves in the hope that they will behave more simply than the nucleons which they so profoundly influence. The final paper in this series will give an account of the preliminary experiments at Berkeley, which have the objective of determining the properties of mesons.

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## Causality, Relativity, and Language

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(Received November 6, 1950)

The development of modern physics requires that verbal interpretations given to the symbols used in mathematical analysis be closely examined to determine whether such verbal interpretations are meaningful or even possible. The concept "causality" is studied. Considerations suggested by relativity theory indicate that this concept can have only a local meaning and that, in particular, two observers in different parts of the universe could not communicate adequately with each other if they employed the concept of causality in their statements of physical phenomena. It is requisite that the scientist give up his insistence on verbal translations of abstract symbols in order that he be able to make more general and profound insights in his analytical reasoning.

GENERALLY speaking, the process of physics is to set up our observations of world phenomena in analytical or mathematical statements. The symbols employed by these statements are defined verbally, in the language derived from the experiences of our own senses. Further, those symbols which appear as a consequence of our analytical reasoning we usually attempt to translate into the verbal language. To be unsuccessful in this attempt, that is, to leave the symbols as untranslatable components of our mathematics or analytics, is to leave them "not satisfactorily explained" or "not fully understood." To be unsuccessful in this attempt is to leave ourselves a certain discomfort, since it is the verbal language by which our psyche expresses itself.

One might suggest that it may often be impossible to translate from our analytics into the verbal language simply because the experience out of which has grown our verbal language is limited. This means only that the verbal language is derived from our direct experience, while the symbolic language is the description of our indirect experience. As there is no necessity that the world we know by our instruments is congruous with the world known to us by our five-sense contacts, there is no necessity that the language we use to describe one or the other will be the same or even mutually translatable.

The phrases, then, "not satisfactorily explained" and "not fully understood" are false descriptions of our state of knowledge. While it may not be for some of us psychologically satisfying if symbols are not translated into the verbal language, we must admit that they are "understood" in that we can employ these symbols to make valid analytical statements about the world. This, may I suggest, is all we can properly define "understanding" to mean.

Accepting what has been stated above, it is important to inquire into the meaning of many of our verbal concepts to see whether they are still applicable and to try to discover any limitations of their application or syntax. It will be the purpose of this paper to study the concept, "causality," particularly in the light of relativity theory.

We understand by the statement "event $A$ is the cause of event $B$" that the physical laws describing the relationship of $A$ to $B$ demonstrate the necessity of event $A$ occurring in order that event $B$ shall occur. In our own thinking, that is to say, in the general verbal use of the concept "causality," it is understood that if $A$ caused $B$, $A$ preceded $B$. Before the advent of the relativity theory we might have been prepared to admit that if $A$, the cause, was communicated to $B$, the effect, with infinite velocity, there would not be any temporal separation of the cause and the effect. We have learned, however, that the velocity of light is the limiting velocity; and hence this threat to our feeling for the necessity of temporal separation of the cause from its effect is abated.

The problem of simultaneous events arises. If the event $A$ appears to us to be simultaneous with the event $B$, what can we say of their causal relation? We seem to have two alternative answers to this question. We may argue that not all events are causally related, even if they are temporally separated, and thereby dismiss the question. Or we may admit that the condition of temporal separation is not necessary to our understanding of causality and seek an understanding of the concept which has nothing to do with time.

Consider first the second of these alternatives. Since in our own minds time has ordered all of our observations and all of our perceptions, it is unreasonable to suppose we can divorce time from our verbal interpretations of these observations; that is we, cannot dismiss time from a concept which is derivative from these observations and perceptions. In other words, our verbal language is the product of our psyche, and we cannot understand or construct words which do not reflect our own psychology. It is meaningless, therefore, to seek an understanding of the concept "causality" which does not imply a temporal ordering of cause and effect as long as we have no experience with the instantaneous communication of the cause with its effect.

In order to deal with the second alternative let us postulate that all events which are temporally separated are causally related. Whether they *can* be so related depends only upon the events chosen and the completeness or comprehensiveness of our knowledge of the events of the world.

To demonstrate the validity of the postulate consider the case of three billiard balls, $a$, $b$, and $c$. Ball $a$ is struck and moves to ball $b$, knocking $b$ into a pocket. Ball $a$ strikes the side of the billiard table and rebounds a number of times, performing what is generally described as random motion. A number of seconds after it hit $b$ ball $a$ strikes ball $c$. According to the postulate, the event "$a$ strikes $b$" may be considered the cause of the event "$a$ strikes $c$." This is clear when we admit that the motion of ball $a$ after striking ball $b$ is, of course, not random but may be described by the laws of mechanics, and when we admit that, had ball $a$ not struck ball $b$, the event "$a$ strikes $c$" would not have occurred.

It is only a matter of convenience that we choose to consider ball $a$ the cause of "$a$ strikes $b$" and of "$a$ strikes $c$," for this consideration may simplify a mathematical analysis of the situation. But theoretically (and in this case actually) we could consider the event "$a$ strikes $b$" the cause of the event "$a$ strikes $c$" with the cause being transmitted from the first event to the second via the ball $a$.

Notice now how this situation would appear if we knew nothing of the intermediary ball $a$ in the time between the event $b$ (described simply as ball $b$ set into motion in a given direction) and the event $c$ (described as ball $c$ set in motion) which succeeded it. Could we with any justification think that there is no causal relation between the events $b$ and $c$? There is a causal relation, and this relation is identified because the events are separated temporally; it only remains for us to discover what the mechanics of the situation are, it only remains for us to learn of the interloper ball $a$.

The implications of the foregoing can be generalized. If we presume (and we do presume) that Nature may be described by laws consequent from our inquiries into the phenomena of the universe, and if we imagine a fundamental cause which figuratively was the beginning of the universe and from which developed in a chain reaction events which became the cause of further events, etc., any of the events on this branch-work which we perceive as temporally

separated may be considered causally related, with one event the cause and the other the effect.

How then can we resolve the problem of the causal relationship of events which appear to us to be simultaneous? Here the relativity theory comes to our aid; for it is a consequence of the Lorentz equations that events which appear to be simultaneous to one observer may not necessarily appear simultaneous to an observer who is in motion relative to the first-mentioned observer. This can be easily seen in the simple case where the first observer is stationary with respect to the events while the second observer is moving with uniform velocity relative to the first observer. If two events occur at times $t_1$ and $t_2$ for the stationary observer such that $t_1 - t_2 = 0$, the time interval between the events for the moving observer will be, from the Lorentz equation for transforming time (motion along the $x$-axis):

$$t_1' - t_2' = \frac{t_2 - t_1 - v(x_1 - x_2)/c^2}{(1 - v^2/c^2)^{\frac{1}{2}}} = \frac{v(x_2 - x_1)/c^2}{(1 - v^2/c^2)^{\frac{1}{2}}},$$

where $v$ is the velocity of the moving observer, $c$ the velocity of light, and $x$ the distance of the events from the origin of the stationary system. It is apparent that the time order of the events as seen by the moving observer will depend upon the direction of his motion with respect to the stationary system.

For the moving observer spoken of above, the events are temporally separated; and hence we can expect that he may relate them causally, choosing for the cause the "first-occurring event." This could be either event, depending upon his direction of motion. Since he presumably would describe the relationship of the events using the same physical laws as we do, one is tempted to ask, "Are they really causally related?" The answer to this might be that, since there is no absolute observer, each person has a right to describe the world as he sees it and to define his view as real, and we may assert that the question is therefore meaningless.[1] That events which for a given observer are not temporally separated cannot by him be causally related can be simply considered a limitation of his perceptions.

The question may be examined a little more deeply, for the inquirer may insist, "There are some laws of physics, e.g., the Maxwell equations for electrodynamic phenomena, which are invariant under relativistic transformations and which may be employed to describe the events in an absolute causal relationship."

The difficulty with this proposition is just that discussed in the first part of this paper; our verbal language is conditioned by experiences which are limited, here limited by the restrictions known from the relativity theory. We cannot consider the symbol $ds^2$, for example, as having any verbal meaning; we cannot have any direct experience with the so-called world of four dimensions and hence we cannot invent words, e.g., "causality," to describe such a world.

We find ourselves isolated observers examining the universe by partitioning space and time according to our peculiar locations. If two observers in different parts of the universe were to communicate with each other about two events, they could not in the case described above speak to each other of the "causal relation" of the events but rather would have to present each other with mathematical statements which each could translate, where possible, into his own verbal language.

We can see that a verbal language is largely a private affair. Only when two observers share similar experiences will they be able to communicate with each other satisfactorily. Certainly, when their perceptions of the world are different, will their communication be hampered.

It is therefore necessary to avoid confusion between the symbolic or analytic language that is used to describe essentially common experience (the body of physics or science) and the verbal language which is employed to discuss essentially private experience. We may expect to communicate among ourselves better as scientists exchanging with each other mathematical statements of the universe than as mere mortals handicapped by the limitations of our peculiar tongues. In particular, it is important for us as scientists to realize when we may be falling into the error of mixing the two different languages; for, once we surrender our demand for the translation of analytical symbols into verbal syntax, we clear the way to making more profound and fruitful insights into abstract mathematical reasoning.

[1] E. Fein, Am. J. Phys. **13**, 296 (1945).

# A Laboratory Course in Electronics

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(Received July 21, 1950)

An electronics laboratory course developed primarily for junior and senior students in physics is described. Emphasis is placed on the theory of operation of vacuum tubes and their uses as circuit elements in research applications. The equipment, largely home-constructed and described in some detail, is provided in a form which allows quick assembly by the student, and individual components are designed to provide maximum flexibility in setups.

During the past few years a laboratory course in electronics has been developed at Yale which may be of general interest. It is designed for juniors and seniors majoring in physics and differs somewhat from the usual course offered to students in electrical engineering. Considerable emphasis is placed on the understanding of physical principles involved as well as the acquiring of techniques. We feel that the student who graduates today whether he goes into industrial research or continues his studies in graduate school needs considerable experience and skill in the use of electronic techniques and that his understanding should be based securely enough on physical principles that he will be able to handle the nonstandard situations which continually arise in physical research rather than the more standard applications of engineering. We do not feel that this course teaches physics but rather a necessary branch of engineering with emphasis on the physical principles and applications.

It has been found convenient to divide the course into two terms, quite different in their approach. In the first term the students follow detailed instructions and have little time for independent exercises, while in the second term they are allowed much more freedom.

## FIRST-TERM EXPERIMENTS

In view of the fact that time seems all too short to train a previously unprepared man in modern electronic techniques, the first-term work consists of a set of ten experiments which the students do in pairs, all at the same time. The first-term course accompanies a lecture course meeting twice a week in which Reich's book[1] is used as the textbook, but generously extended with regard to physical principles and as generously cut with regard to engineering applications.

Since in the first week the students have not learned enough to do a respectable experiment, the first laboratory period is very brief, serving only to introduce them to the methods of setting up the experiments and using the power supplies. The first experiment is on the study of diode characteristics. The students use milliammeters and a vacuum-tube voltmeter together with the filament voltmeter on the power supply to obtain plate current *vs* plate voltage curves with different values of heater voltage for 6H6, 6X5, 2X2, and 3B24 tubes as examples of very closely spaced and widely spaced diodes. They also check the empirical relation $i_b = ke_b^n$, where $i_b$ is the plate current and $e_b$ the plate potential, and determine the value of $n$ from their data. If they have time, they obtain a diode characteristic with a series load resistance.

In the second experiment the characteristic curves of the 6C5 triode are determined, $i_b$ *vs* $e_b$ with the grid potential $e_c$ as a parameter and $i_b$ *vs* $e_c$ with $e_b$ as a parameter. Then, from these same data, $e_b$ *vs* $e_c$ with $i_b$ as a parameter is plotted. The slopes of these curves give the value of $\mu$ (which is also measured with the home-made tube coefficient bridge described below) as well as the values of $g_m$ and $r_p$. In the next experiment a similar study is made of the characteristics of the pentode tubes 6SJ7 as a pentode and as a tetrode and the 6SK7 as a pentode. This requires an auxiliary power supply for the screen.

In Experiment IV, the students study the single-stage voltage amplifier using a 6C5 tube. The amplifier is wired by the students with spade-tipped leads on the amplifier board. All the major items are mounted and terminals are provided

[1] Herbert J. Reich, *Theory and Applications of Electron Tubes* (McGraw-Hill Book Company, Inc., New York, 1944), second edition.

FIG. 1. A view of the equipment used to study the characteristics of a two-stage voltage amplifier. In the background may be seen the homemade audio oscillator and voltage-stabilized power supply and cathode-ray oscilloscope, and in the foreground, the vacuum-tube voltmeter, audio attenuator, and amplifier board.

for the cathode resistor and by-pass, plate load resistor, and other small components. Using a homemade audio oscillator, homemade attenuator, Heath-kit vacuum-tube voltmeter and oscilloscope, the students observe distortion, measure voltage amplification as a function of load resistance, plot output voltage *vs* input voltage, measure the effect of the cathode by-pass condenser, and study the effect of an ac load resistance which is low compared to the dc resistance in the plate circuit. Then they reconnect the amplifier as a cathode follower and determine its voltage amplification, gain, and output impedance.

Experiment V is a similar study of a two-stage voltage amplifier, but the oscilloscope is used to measure phase shift so that voltage amplification *vs* frequency and phase shift *vs* frequency curves are obtained. In this case the students calculate the proper value of cathode resistor and study the effect of different values of coupling condensers and second-grid resistors. Provision is made in the board for insertion of negative feedback and its effect is measured. The setup used for this experiment may be seen in Fig. 1, and the circuit in Fig. 2.

For the sixth experiment on class $A$ and $AB$ power amplifiers, in which the students measure power output *vs* load resistance with constant excitation and then so-called "optimum" power output (i.e., maximum power output consistent with a given amount of harmonic distortion), it was necessary to construct a simple single-frequency filter for the input and a simple single-frequency harmonic analyzer. The circuit of the experiment is shown in Fig. 3. The filter ensures that there is practically no harmonic distortion in the input signal. The microammeter (a cheap

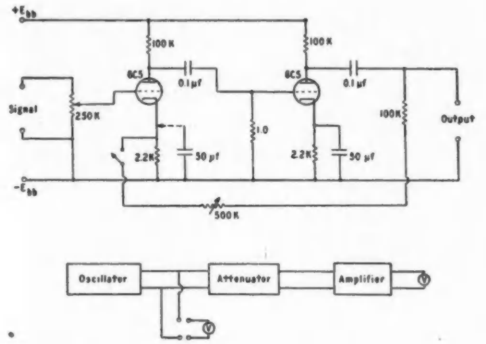



FIG. 2. The circuit of the two-stage voltage amplifier studied and a block diagram of the experimental arrangement.

one) is used to detect the point at which the grid is driven positive. The circuits of the harmonic analyzer and filter are shown in Fig. 4. The switch allows one to connect either to the second harmonic which is passed by the analyzer filter or to the total output (essentially the fundamental for small distortions) through a twenty-to-one potential divider. With this arrangement, if there is no change of the voltmeter reading on switching, the harmonic distortion is five percent (since harmonic distortion is defined as the ratio of harmonic to fundamental *voltage*, not power). Also, if the student realizes it, he may arrange it so that the voltmeter reads five on some scale when switched to the fundamental in which case the voltmeter will read the harmonic distortion in percent directly when the switch is thrown. A set of curves obtained by a student is given in Fig. 5. The experiment ends with a brief examination of a push-pull amplifier with triodes and then pentodes, using the same tubes.

The student is now considered to be thoroughly familiar with amplifier characteristics and the next experiment is concerned with amplitude

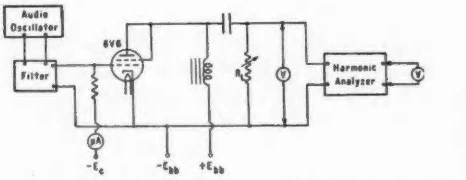



FIG. 3. The circuit which is used in the study of the power output and distortion of class *A* and class *AB* power amplifiers.

modulation. This is a rather qualitative experiment in which the modulation characteristic of a linear plate modulator is measured using a carrier frequency between 20 and 50 kilocycles per second and modulating it at the power line frequency. A diode detector is made up and demodulation observed. All the waveforms are observable with the Heath-kit oscilloscope. If time permits, the student sets up and studies a linear grid modulator.

The two-stage amplifier board is used in Experiment VIII to construct an Eccles-Jordan trigger circuit and then a multivibrator circuit. The students vary the values of the components and observe the behavior of the circuits and make a thorough study of the wave forms involved. They next set up a differentiating circuit

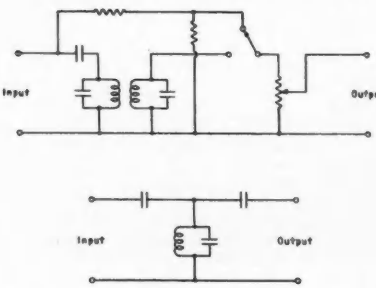



FIG. 4. The circuits of the homemade harmonic analyzer (above) and 1500-cps band-pass filter (below) which are used in the power-amplifier experiment. The harmonic analyzer is provided with a switch to allow either fundamental (plus second harmonic) or second harmonic output to be obtained.

and an integrating circuit and observe the effects which these circuits produce on the pulses derived from the multivibrator circuit. Then they experiment with frequency division by driving the multivibrator with a higher frequency from the audio oscillator.

Glow and arc-discharge tubes are studied next by plotting the current *vs* voltage characteristic of an OC3/VR105 tube with both normal and reversed connections. The ignition and extinction voltages are determined and then a relaxation oscillator is set up and observed. The grid control characteristic of a 2050 tube is plotted and then the tube is used as a grid-controlled rectifier in the circuit shown in Fig. 6.

In the tenth experiment, the static current-voltage characteristics of a type 929 vacuum

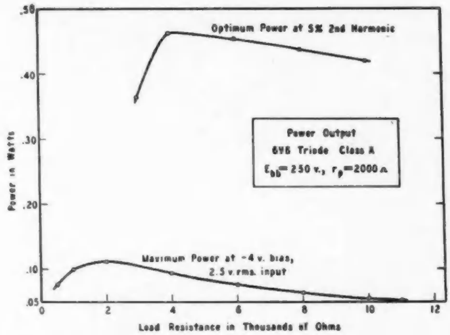



FIG. 5. A student's set of curves for the power amplifier experiment which show that at a constant signal level the greatest power output is obtained when the tube and load resistances are equal but if the signal is changed to obtain five percent harmonic distortion greatest power occurs for twice that load resistance.

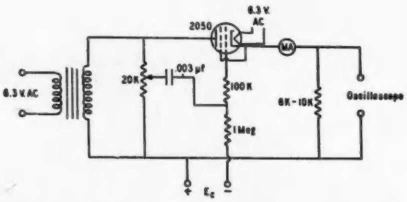



FIG. 6. A circuit in which a 2050 tube is used as a grid-controlled rectifier as part of the experiment on glow and arc-discharge tubes.

phototube and a 918 gas phototube are measured and then, with a lamp bulb in a black-painted box, the spacing of such curves as a function of the relative light flux (using the inverse square law) are studied. The currents (which are less than 20 microamperes) are measured with a series resistance and the vacuum-tube voltmeter. By using a neon lamp on the 60-cycle line as a modulated light source, the modulated output of the photo-tube may be observed.

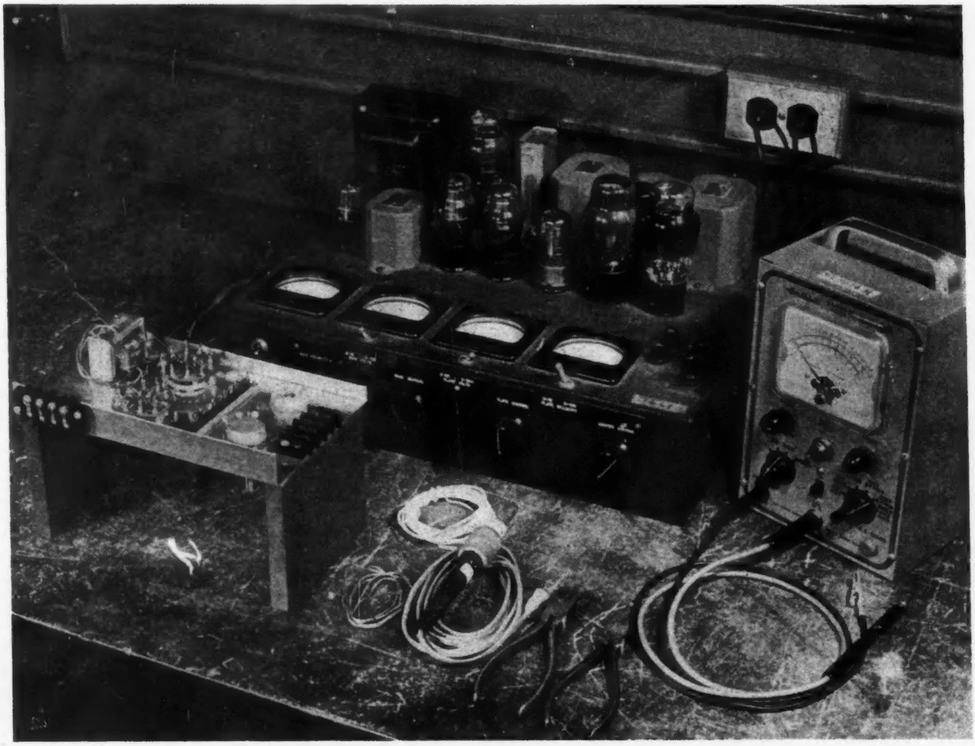


FIG. 7. A view of the equipment supplied the last session of the first term with which the students construct and test voltage stabilizer circuits. This is the first experiment in which the students are allowed to construct equipment.

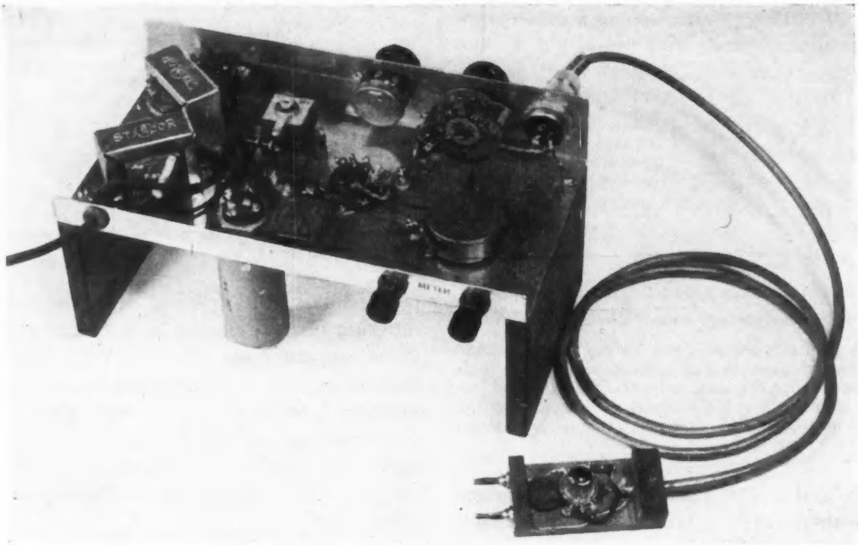


FIG. 8. The chassis for the second-term experiment on vacuum-tube voltmeters. The students receive it as shown so that their mechanical work is confined to wiring beyond the power supply and filament circuits. This chassis is typical of those used for student-constructed circuits.

For the last afternoon of the course, instead of doing an experiment of the type which they have been doing, the students are given the mounted chassis shown in Fig. 7, a small Ungar soldering pencil, wire, parts, and solder. The previous week they have been instructed to look up Hunt and Hickman's paper on voltage stabilizers[2] and so are able to put together quite a respectable example of a voltage stabilizer, whose stabilization ratio and internal resistance they measure. If they have time, they make the usual slight variations on the circuit and observe the effects on the stabilization ratio and internal resistance.

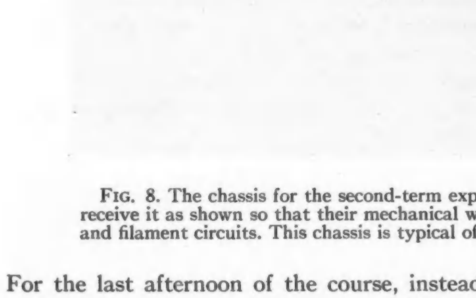



FIG. 9. The circuit which the students construct for the study of vacuum-tube voltmeters. The probe is supplied complete as may be seen in Fig. 8.

## SECOND-TERM EXPERIMENTS

In the second term there is no accompanying lecture course and the students are sufficiently experienced that they can do advanced experiments relatively independently. Only one setup of each experiment is provided and the students work in pairs shifting from one experiment to another at their own speed. In each of the seven experiments they are required to wire the circuits, decide upon component values, and then make the circuit work. Before starting the experiment the students are given references and sent to the library to study the literature on the particular device concerned.

Figure 8 shows the chassis of the vacuum-tube voltmeter experiment as the students receive it and Fig. 9 shows the circuit used. The probe is assembled and wired for them since its construction is somewhat delicate and critical. They adjust the circuit, calibrate it on all scales at 1000 cycles/sec and at the highest frequency

[2] F. V. Hunt and R. W. Hickman, Rev. Sci. Instr. 10, 6 (1939).

readily available in the laboratory (100 kc/sec or so). Then, omitting the probe, they calibrate the voltmeter on direct current, change the circuit to make a plate-detection ac voltmeter, and any do additional experiments which occur to them.

In the special wave form-generation experiment a blocking oscillator is constructed, triggered from a synchroscope which is used to study the output pulses. Next they build a triggered single-sweep sawtooth generator and study its behavior with various changes of components and methods of insuring a linear sweep. Then they make a pulse-forming network with a thyratron to discharge it and study the waveforms involved. The circuits are shown in Fig. 10.

To study electronic computing circuits, the students build and work with the difference amplifier, adding circuit, and squaring circuit shown in Fig. 11. If they wish, they work on other computing circuits which they find in the literature.

To study the resistance-capacitance oscillator a chassis similar to that shown in Fig. 8 is used

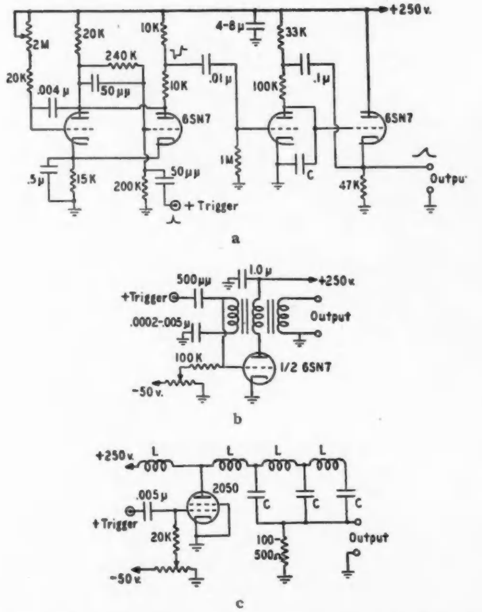



FIG. 10. Special wave forms are generated and studied with the circuits shown. (a) Typical wave forms are shown at the appropriate points of the sawtooth generator circuit. (b) Blocking oscillator. (c) Square-pulse generator.

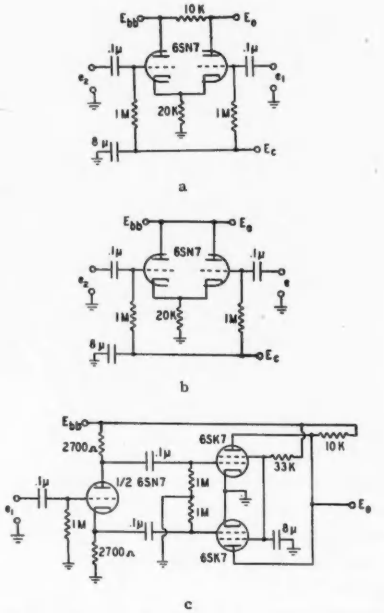



FIG. 11. Three circuits used to investigate the principles of electronic computers. Students frequently make variations on these circuits in the course of their experiment. (a) Difference amplifier. (b) Adding circuit (c) Squaring circuit.

to build the circuit of Fig. 12. The values of the components for the phase-shifting network are computed by the students who are required to explain in their report the functions of each component in the circuit and the reason for the use of the two cathode followers.

There are two experiments on amplifiers, one on the video amplifier whose chassis is shown in Fig. 13 and one on 27 mc/sec tuned amplifiers. The students measure the characteristics of the video amplifier without peaking coils and then wind shunt and series peaking coils and experiment with them. The tuned amplifier, whose circuit is shown in Fig. 14, is studied with the aid of a Heath-kit frequency-modulated signal generator. It was found necessary to modify this generator to obtain satisfactory results. The main tuning shaft was shielded to prevent rf leakage, a 470-ohm resistor was inserted in series with the marker frequency input jack to prevent interaction with the external oscillator, and it was found necessary to distort greatly the tuning condenser plates to obtain agreement with the

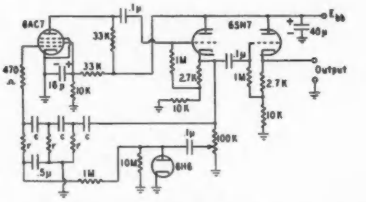



FIG. 12. The resistance-capacitance-tuned oscillator circuit in which the students design the RC network as well as construct the circuit and make it work.

frequency scale. After measuring the amplifier band width and comparing it with the calculated value, a similar double-tuned amplifier is studied.

The last experiment is on the quartz-crystal oscillator. War-surplus crystals are given the students to regrind to a specified frequency. They make simple crystal oscillator circuits and study their behavior. In this way they also become familiar with precision frequency measurements.

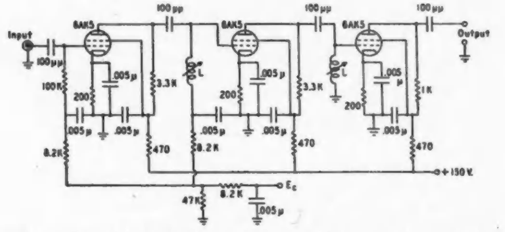



FIG. 14. This single-tuned amplifier and a similar double-tuned amplifier both with a center frequency of 27 Mc/sec are wired, adjusted, and measured using a frequency-modulated signal generator, with oscilloscopic display of the frequency characteristics.

### HOME-MADE EQUIPMENT

As may be seen from the illustrations above, the "boards" are simple open boxes with a bakelite plate over one end on which all the parts are mounted. They are made so that they nest for easy stacking in the storage cabinets. The nickelled binding posts are spaced so that General Radio shorting plugs may be used for making connections wherever posible. The aim

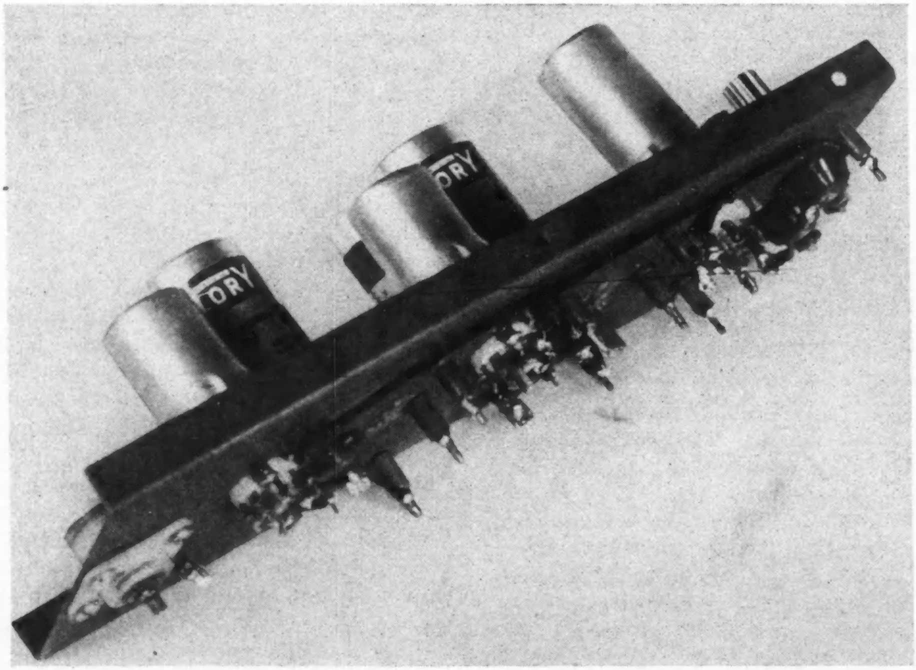


FIG. 13. The unwired chassis of the video amplifier as supplied the students. This is typical of the construction used for high frequency circuits in this course and is both an example of good engineering practice and readily accessible for wiring by the students.

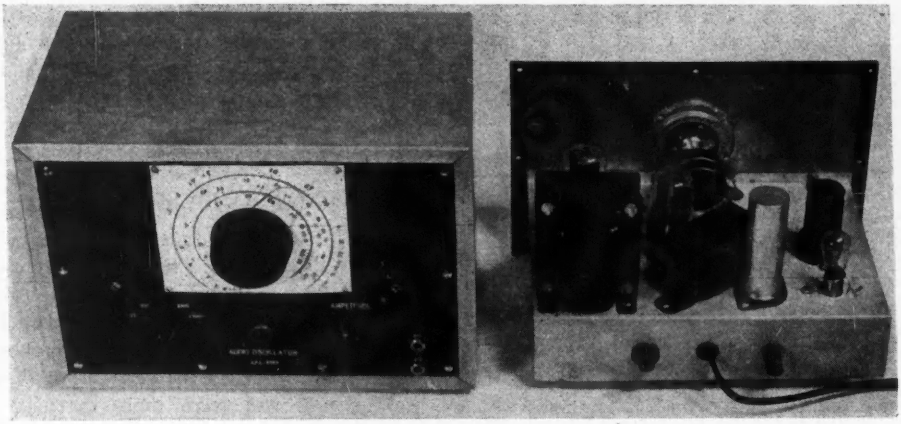


FIG. 15. This audio oscillator using a Wien-bridge circuit is typical of the homemade instruments constructed for the course.

is to minimize the time the first-term students spend in mechanical construction. Connections between pieces of equipment are made with spade-tipped leads.

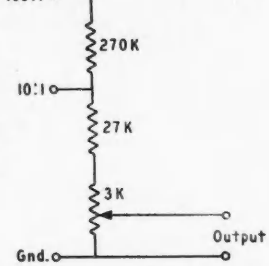
Separate electronically regulated power supplies are provided for each pair of students. One of these power supplies may be seen in Fig. 1. Many parts were war surplus. The supplies provide 6.3 volts ac with a series rheostat and a voltmeter, 0 to 150 volts for grid bias with a voltmeter and a switch to change the polarity with respect to ground, and 0 to 300 volts electronically stabilized with a voltmeter and a milliammeter which, were we building them now, we would omit because an external home-made multi-range milliammeter has been found more useful. All three power sources are fused and it is well to have the fuses readily replaceable.

The audio oscillator is a lamp-stabilized Wien bridge oscillator[3] with an added buffer amplifier. Its style of construction may be seen in Fig. 15. The circuit of the audio attenuator is shown in Fig. 16. It is actually a potential divider since it is to be used in a high impedance circuit. The output potentiometer has a low value of 3000 ohms so that any shunting capacitance of the following-amplifier input will be negligible. In this case the division ratios of the attenuator are always the same. The potentiometer is provided with a scale calibrated evenly from 0 to 1.0.

Two tube-coefficient bridges[4] were made for use in the second first-term experiment with the circuit shown in Fig. 17. The construction is shown in Fig. 18.

The simplicity of the homemade harmonic analyzer is made possible by the fact that for a triode the harmonic distortion is practically all even and the fourth harmonic term is small. It is sufficient to filter out only the second harmonic of the input signal frequency and consider that as the total harmonic content of amplifier output. As may be seen from Fig. 4, the filter for the oscillator output is a simple type tuned to about 1500 cycles/sec. The harmonic analyzer uses a double-tuned transformer tuned to about 3000 cycles/sec. The two circuits are slightly over-

FIG. 16. The circuit of the simple potential divider which is used as an audio attenuator. This is permissible because of the high input impedance of the circuits with which it is used.




[3] Herbert J. Reich, *Theory and Applications of Electron Tubes* (McGraw-Hill Book Company, Inc., New York, 1944), second edition, p. 397.

[4] E. H. Schulz and L. T. Anderson, *Experiments in Electronics and Communication Engineering* (Harper and Brothers, New York, 1943), p. 137.

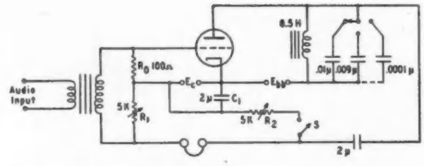



FIG. 17. The circuit of the homemade tube coefficient bridge. Balances are obtained with $S$ open and closed. The students are required to prove that:

$$\mu = R_1/R_0 \text{ at balance, } S \text{ open;}$$
$$r_p = (\mu R_0 R_2/R_1) - R_2 \text{ at balance, } S \text{ closed;}$$

and

$$g_m \cong R_0 R_1/R_2 \text{ at balance, } S \text{ closed, if } R_2 \ll r_p.$$

coupled to give a flat pass-band to allow some leeway in setting the frequency of the oscillator. There is a voltage step-up in the transformer of about three-to-one due to the resonance. For this reason and because of the loss in the pass-band of the network the potential divider for the fundamental is not actually twenty-to-one as it appears but is empirically adjusted to give that effect.

We have found that these simple homemade instruments have been quite adequate for the purpose and that their cost has been kept down by keeping our eyes open for war-surplus parts and by the use of laboratory labor. At the beginning of the course we also constructed oscilloscopes and vacuum-tube voltmeters, but this is no longer necessary.

## PURCHASED EQUIPMENT

The Heath-kit[5] vacuum-tube voltmeters have been very satisfactory. The parts are of excellent quality and the cases, although light in construction, have stood up well under student use. The oscilloscopes have been satisfactory except for slight modifications to the vertical amplifiers.

For the second term more extensive commercial equipment is necessary. The tuned amplifier experiment requires a frequency-modulated signal generator. A crystal diode probe, either home-made or kit, in conjunction with the Heath-kit vacuum-tube voltmeter, is needed for the video amplifier experiment as well as a 10 kc/sec to 30 Mc/sec calibrated signal generator and a $Q$-meter (an inductance bridge could be

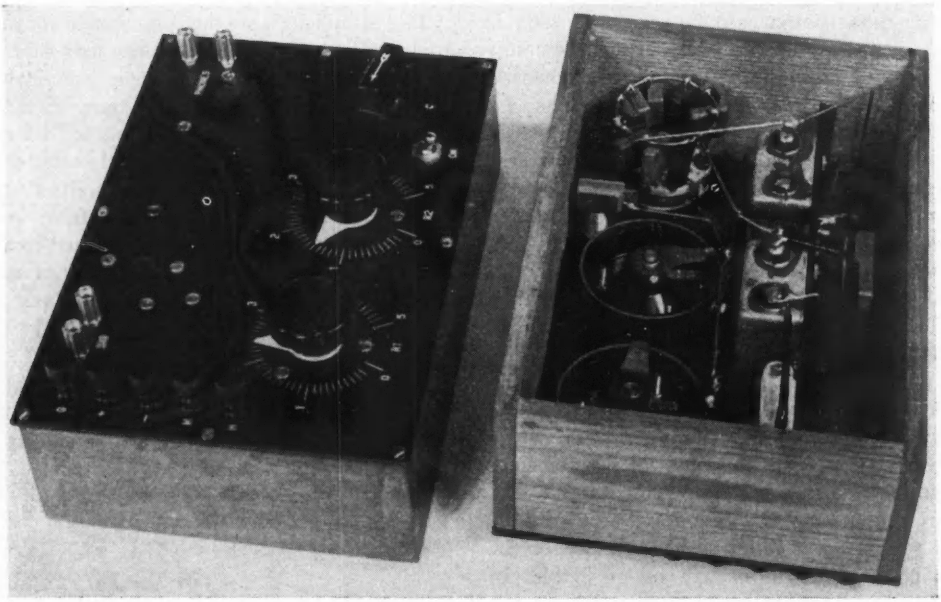


FIG. 18. Views of the two homemade tube coefficient bridges constructed of war-surplus parts in the style of the low frequency "boards."

[5] Manufactured by the Heath Company, Benton Harbor, Michigan.

used instead). The $Q$-meter and a heterodyne frequency meter are used in the crystal oscillator experiment. A synchroscope with sweep ranges of 5, 50, and 1000 microseconds is necessary for the waveform generator experiment.

### RESULTS

We have found it possible to develop a thorough laboratory course in electronics at much less cost than would be thought possible at first glance by building a great deal of the equipment ourselves and making use of the excellent instrument kits available. At all times the aim has been to make as economical use of the students' time as possible in order to teach them enough to make them competent to handle the problems which will arise in their graduate research or in their positions as research assistants.

That this has been accomplished has been proven by the experience that we and other members of the department have had with men who have taken this course. They have been found to have a considerable degree of competence in the design and construction of electronic equipment for physical research and are sufficiently familiar with the literature that they can dig out information on any new device with which they are confronted.

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# Some Behavior Objectives for Laboratory Instruction

Haym Kruglak
*University of Minnesota, Minneapolis, Minnesota*
(Received October 31, 1950)

A sample outline of objectives illustrates a method of laboratory instruction in elementary electricity in terms of student behavior. The twenty-six specific objectives are grouped under five categories: instrumental skills, skills in the use of the controlled experiment, problem-solving skills, miscellaneous skills, functional understanding of principles and habits.

A LARGE number of virtues have been ascribed to laboratory instruction in science. But though there might be some agreement on the most general, traditional objectives, the problem of measuring the outcomes of laboratory teaching is far from formulated, let alone solved.

A questionnaire submitted in 1926 by Hurd to science teachers at the University of Minnesota illustrates "the variety and confusion of opinion as to objectives in laboratory work" (p. 8).[1] Thirty-five respondents listed forty-three statements on the function of laboratory work in science, with no more than eight individuals in agreement on any single objective. Nor is there much evidence that laboratory instruction at the present time is being evaluated effectively in terms of specific behavior objectives. A recent personal experience will perhaps illustrate the latter point.

A few months ago the writer attended a national conference on science education. Much learned and often heated discussion went on among the delegates, but no decisions were reached and no resolutions were adopted. After one of the sessions, one of the country's leading science teachers turned with a worried expression to the writer and said: "What do we teach laboratory for, anyway?"

The pedagogical importance of formulating objectives has been discussed by Nedelsky[2] in an elegant, thorough, and cogent manner. His argument for a two-dimensional list of objectives re-emphasizes the fact that every pedagogical process implies subject matter *and* students.

The list of behavior objectives for laboratory instruction in elementary physics shown below has been developed as a guide for the construction of laboratory performance tests. Professor Nedelsky prepared a very useful manual with test items designed to illustrate the testing for specified objectives in physics.[3] However, the

[1] A. W. Hurd, *Problems of Science Teaching at the College Level* (University of Minnesota Press, Minneapolis, 1929).

[2] Leo Nedelsky, Am. J. Phys. **17**, 346–54 (1949).

[3] Leo Nedelsky, *Testing for Specified Objectives of Physics Teaching* (The College and Examiner's Office, University of Chicago, Chicago, Illinois).

manual contains examples of paper-and-pencil test items only.

### OBJECTIVES OF LABORATORY INSTRUCTION IN ELEMENTARY ELECTRICITY

I. Instrumental Skills

A. To manipulate basic apparatus

1. To wire a simple circuit
2. To balance a bridge circuit
3. To adjust a galvanometer telescope and scale
4. To make a good electrical connection to a terminal post

B. To make apparatus setups from memory

1. To set up a circuit for measuring resistance by the voltmeter-ammeter method
2. To set up a slide wire Wheatstone bridge

C. To make basic measurements with given apparatus already set up

1. Resistance
2. Current
3. Potential difference
4. EMF
5. Power
6. Field strength
7. Capacitance

D. To identify various pieces of laboratory equipment and to recognize the function of each

1. Rheostat
2. Decade boxes
3. Standard cell
4. Wire potentiometer
5. Fixed resistors
6. Condenser
7. Lamp bank
8. Potentiometer, K-type
9. Transformer
10. D'Arsonval wall galvanometer
11. Ballistic galvanometer
12. Littel fuse
13. Galvanometer shunt
14. Inductance
15. Standard resistance
16. Wheatstone bridge (2 types)

E. To trace connected circuits for accuracy and to make conventional diagrams from connected circuits

1. Given a connected circuit containing ammeter, rheostat, voltmeter, resistance, switch, and battery

a. Is the circuit connected correctly?
b. Draw a conventional schematic diagram.

F. To read and interpret graphs, tables, and charts

1. To read a wavemeter calibration graph
2. To use a copper wire table
3. To interpret tube characteristics

G. To develop accuracy in measurement

1. To read an instrument with a mirror scale
2. To compute error from experimental data
3. To find the zero reading of an instrument and to know how to apply the correction
4. To be able to make an error analysis

II. Skills in the use of the controlled experiment

A. To recognize the adequacy or inadequacy of controls

1. Triode board without filament ammeter
What instrument should be used to be sure that the filament temperature is constant?

B. To recognize the appropriate experimental procedure for the measurement of a quantity

1. A circuit is set up. A sequence of operations is described. Will data obtained yield the required quantity?

C. Given a setup and an objective, to determine whether one or more parts is missing to reach the objective, and to name those parts

1. Joule's law setup without clock or thermometer

D. Given miscellaneous apparatus, to select the necessary pieces to reach an objective

1. To select the pieces needed to calibrate an ammeter by the electrolysis method

E. Given a piece of equipment, to determine whether it is in working condition

1. Resistance box with one or more coils open. Locate the open coils with a dial Wheatstone bridge.

III. Problem-solving skills

A. To solve problems, new to the student, involving the use of laboratory apparatus

1. Given a blackened bulb and other apparatus, find whether the bulb has a carbon or a tungsten filament.

B. To solve problems based on data obtained in the laboratory

1. Given laboratory data on plate characteristics of a triode, find the amplification factor of a tube.

IV. Miscellaneous skills

A. Ability to recognize an unsafe circuit

1. To recognize a short circuit
2. Given a circuit with ammeter in parallel and voltmeter polarity reversed, to determine what is wrong

B. Ability to remedy unsafe circuits by modification

1. Remedy IV A 1 above
2. Remedy IV A 2 above

C. Ability to follow directions in a laboratory situation

1. Oral directions
2. Written directions
3. Directions by means of diagrams

a. Given a circuit diagram and components, connect as per diagram

D. Ability to arrange data in tabular form

E. Ability to make a good graph

F. Ability to write a concise, effective report

V. Functional understanding of principles

A. To predict, on the basis of theory, what is likely to happen with a given laboratory setup

1. Given galvanometer with external wire shunt, connected as an ammeter. What will happen to the deflection if the wire is shortened?

B. To recognize the basic principle which is used in a given laboratory setup

1. Given an earth inductor, how should it be placed to get maximum emf for the vertical component of the earth's magnetic field? Why?

C. To recognize the generalization which is being verified

1. Given several resistors in series with a battery, voltmeter, switch, what generalization can be established with the setup?

VI. Habits

A. Neatness

1. Cleaning up own laboratory station

B. Caution

1. Rechecking setting of potentiometer with standard cell

C. Safety

1. Recognizing and remedying procedures unsafe for operator and apparatus

Professor C. N. Wall and the writer feel that a paper-and-pencil test is seldom, if ever, an adequate substitute for an examination involving the use of apparatus and materials. For the past several years "practical," or performance laboratory tests have been given to all of the students in the elementary physics courses at the University of Minnesota. The items in these tests are in general correlated with the objectives listed above. The results of the testing program are now being collated and analyzed. The description of the test items, the method of administration, the item analysis, etc., will be submitted for publication at some future date.

The writer wished to express his appreciation to Professor C. N. Wall, of the University of Minnesota, for stimulating discussion and fruitful suggestions; to Professor Leo Nedelsky, of the University of Chicago; to Professor T. A. Rouse, of the University of Wisconsin; and to Professor R. Schlegel, of Michigan State College, for critical comments and valuable suggestions.

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*Unprotected, a man could not safely come within a quarter mile of such a nuclear chain reaction as takes place in a Hanford pile. The pile is so shielded, however, that workers move about freely right next door to the storm of radiation.*—Eighth Semiannual Report of the Atomic Energy Commission.

# Cosmic-Ray Neutrons

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*New York University, University Heights, New York, New York*
(Received September 17, 1950)

The discovery of neutrons produced by the cosmic radiation is reviewed, and the evidence today available is summarized. The life history of a neutron in the atmosphere is discussed, as is its eventual capture to form radiocarbon. The latitude effect and altitude dependence are presented. Some of the possible production-processes are discussed.

SHORTLY after the original discovery of the neutron, in the early 1930's, investigators began to wonder whether there were any neutrons present in, or associated with, the cosmic radiation. The first paper to appear in the literature giving definite evidence on this subject was one by L. H. Rumbaugh and G. L. Locher.[1] They sent photographic plates aloft in the famous stratosphere flight of the Explorer II, the balloon flown by Captain A. W. Stevens which reached the record altitude of 72,400 ft on November 11, 1935. These observers studied the tracks in the emulsions, and concluded that a large number of these tracks had been formed by recoil protons which had been struck by fast neutrons. Very little attention has been paid to this paper, and it is not usually included in the neutron literature; but the conclusions drawn by the authors, to the effect that there were a large number of fast neutrons in the upper atmosphere, were undoubtedly correct. The next experiment was that conducted by Fünfer,[2] who operated a boron-lined argon-filled ionization chamber at various elevations up to the top of the Zugspitze (8600 ft) in Austria. He too found a counting rate which he attributed to neutrons, although his chamber detected preferentially the slow ones as contrasted to the fast ones detected with plates.

The terminology used in describing neutron speeds differs from that used for protons. A slow neutron is a neutron whose velocity is that characteristic of an energy of a few electron volts or less, down to thermal speeds. A proton in this same energy range is unobservable except in special apparatus; in most circumstances, of course, it has captured an electron and is simply a hydrogen atom. A slow proton is a proton whose velocity is appreciably below the velocity of light, i.e., a nonrelativistic one. Thus, for example, a million-volt proton is called a slow proton, but a neutron traveling at this same speed and with this same energy is called a fast neutron.

After Fünfer's paper, quite a number of experimenters sought to extend our knowledge of neutrons. Various techniques were used, and experiments were carried out at different elevations. Using boron trifluoride-filled proportional counters, developed for the first time with this application in mind, the author[3] conducted a series of high altitude balloon flights, in which the neutron counting rates were transmitted by automatic short-wave radio to a receiving station on the ground. In these original experiments, the presence of fast neutrons was also established by the use of recoil counters. At the same time, the Montgomerys[4] determined the rate of production of the neutrons at sea level. The induced radioactivity produced by neutrons was used as another detection principle by von Halban, Kowarski, and Magat,[5] who carried a bottle of ethyl bromide in an airplane to an elevation of about 30,000 ft over Paris, and measured the amount of activity in the bromine by the method of isotope separation. Other experiments at about this time helped build up our knowledge.[6]

[1] L. H. Rumbaugh and G. L. Locher, Phys. Rev. **44**, 855 (1936).
[2] E. Fünfer, Z. Physik **111**, 351 (1938).
[3] S. A. Korff and W. E. Danforth, Phys. Rev. **55**, 980 (1939). S. A. Korff, Phys. Rev. **56**, 1241 (1939).
[4] C. G. Montgomery and D. D. Montgomery, Phys. Rev. **56**, 10 (1939).
[5] Halban, Kowarski and Magat, Compt. rend. **208**, 572 (1939).
[6] E. Schopper, Naturwissen. **25**, 557 (1937); E. Schopper and L. Schopper, Physik. Z. **40**, 22 (1939); Heitler, Powell, and Fertel, Nature **144**, 283 (1939); S. A. Korff, Terr. Mag. **45**, 133 (1940); Phys. Rev. **59**, 214 (1941); J. Franklin Inst. **229**, 664 and 802 (1940); Phys. Rev. **59**, 949 (1941).

In the meantime, advances occurred on the theoretical front, and the interpretation of neutron measurements was discussed by Bethe, Korff, and Placzek.[7] The point of view developed by these authors can be briefly summarized as follows: We start by assuming that neutrons are produced, in some manner, possibly by nuclear evaporations. These neutrons, coming from nuclei, will have energies of the order of the nuclear binding energies, a few tens of Mev. Whatever their origin, and we shall discuss this further below, let us next consider what will happen to a neutron with a few Mev of energy, free in the atmosphere. Here it will collide with nitrogen and less frequently with oxygen nuclei. At energies of a few tens of Mev, these collisions will be inelastic, and the neutron will rapidly lose its energy to the target nuclei. The nuclei will absorb the energy and will be knocked into excited states from which they will most often return to the ground state by radiation. Since the lowest level in the nitrogen nucleus is at about 4 Mev, and in oxygen at about 6, the neutrons will after one or two collisions be reduced in energy to something less than 4 Mev. After this, they will continue making collisions, but the collisions will be elastic. By ordinary billiard-ball mechanics, we may compute the history of a nucleus of unit mass colliding with other nuclei fourteen or sixteen times as massive. The neutron will lose some energy on each collision, and it will shortly find itself slowed down and tending to come into thermal equilibrium with the heavier particles.

It might be supposed that this process would result in the accumulation of thermal neutrons, and indeed it would, but for the effects of absorption. Nitrogen has an appreciable capture cross section for thermal neutrons, and this cross section varies, as often happens, inversely with the velocity of the neutrons. Thus, the slower a neutron becomes, the more probable it is that it will be captured; and indeed it will, on the average, be captured before it has slowed down to thermal speeds. On the basis of the latest cross-section data, Davis has computed that half the neutrons attain 0.37 ev, and that 0.04 of them become thermal. Oxygen does not play an important role, for the capture cross section in oxygen is very much smaller than in nitrogen.

The result of this process is that the bulk of the energy originally possessed by the neutron has been dissipated in the gas as ionization, this being the manner in which recoiling nuclei are slowed down and stopped. There will be almost as many $N^{15}$ nuclei formed as there are neutrons. But the $N^{15}$ nucleus is unstable, and decays, emitting a proton, into $C^{14}$. The proton is slowed down, is most probably captured by oxygen, and becomes a part of the stable gases in the atmosphere. The $C^{14}$ is a long period beta-emitter, decaying back into $N^{14}$, so that the total amount of nitrogen in the world is not altered by this process.

However, a free carbon atom in the atmosphere will have a high probability of being captured by an oxygen molecule, forming carbon dioxide. Now carbon dioxide, being heavier than nitrogen and oxygen, tends to sink in the atmosphere, and will eventually either be inspired by plants, or will be washed out of the air by rain as carbonic acid, and so will find its way into the oceans of the world. In the oceans, currents carry individual molecules great distances, and many complex kinds of carbonates are formed. Thus the ocean serves as a reservoir, called by Libby[8] a "diluent reservoir," in which mixing occurs, and the result is a uniform distribution, all over the world, of $C^{14}$, radiocarbon. Some of this water re-evaporates, some is blown about as spray carrying with it some carbonates. The water and carbonates come down again as rain and plants will absorb part of this. Thus, a fraction of the radiocarbon will find its way into the organic material of the world. And since radiocarbon has a long lifetime, $5720 \pm 50$ years,[8] there will be time for complete mixing to occur. Since this radiocarbon is a beta-emitter, it is possible to detect extremely minute amounts of it in various substances.

W. F. Libby[8] has developed special counters for making accurate determinations of the amount of radiocarbon in a given sample. If radiocarbon is uniformly distributed, and Libby has shown that it is by measuring samples from

[7] Bethe, Korff, and Placzek, Phys. Rev. **57**, 573 (1940).

[8] Libby, Anderson, and Arnold, Science **109**, 227 (1949). Engelkemeir, Harnill, Inghram, and Libby, Phys. Rev. **75**, 1825 (1949).

many places, then the amount of radiocarbon in a sample of any material is an indicator of the age of the material, or of the time interval working backwards between the present and the time when the sample last was in equilibrium with the carbon-oxygen exchange in the world. For example, while a tree is growing, it absorbs carbon dioxide; and this process terminates when it is cut for lumber. The radiocarbon content is a measure of the length of time since the wood sample was cut. Thus cosmic rays and nuclear physics have provided a totally unexpected tool for the archaeologist, and have enabled him to date wood from the tombs of Egypt and from the pre-Columbian dwellings in the Americas.

Returning again to the study of the neutrons in the atmosphere, we may first ask whether the energy distributions, to be expected from the arguments cited above, have been checked experimentally. The answer to this is in the affirmative. Energy distributions of neutrons are measurable by using a set of shields, each with a different velocity-dependent absorption spectrum, around the neutron counter. Thus, for example, cadmium has the property of being a strong absorber for low energy neutrons, and not an absorber for high energy ones. For example, a millimeter-thick sheet of cadmium will be quite opaque to thermal neutrons, but transparent to neutrons of energy of a few volts or more. Such a sheet of cadmium will have a fifty percent transmission point at about 0.4 volt, but neutrons of a volt will be absorbed only to a few percent. Hence, a neutron detector operated with and without a cadmium shield will permit a determination of the fraction of neutrons which lie below a few tenths of an electron-volt of energy. Other substances have different absorbing properties.

In a series of balloon flights, Korff, Hammermesh, George, and Kerr[9] determined the counting rates of neutron counters surrounded by boron and cadmium shields. The shields were so arranged that they would be in turn slipped over the counter and then removed, in a cyclic operation, while the instrument was in flight. The observed counting rates were transmitted by short-wave radio to the ground station. The counting rates, with and without the several shields, were measured up to 66,000 ft, or through nine-tenths of the atmosphere. The observations were found to support the energy distributions computed by Bethe, Korff, and Placzek.

We may next consider the processes by which the neutrons were originally formed. We may say at the start that, because the neutron is believed to be radioactive, with a half-life of the order of twenty minutes, the neutrons could not be primary particles, coming from any point in space further than twenty light-minutes distant, a distance comparable to the radius of the orbit of the planet Jupiter, and hence a clear proof of the local character of the neutrons. Presumably, most of them are produced as secondary events in the atmosphere, as a result of the incidence of the primary cosmic rays themselves, or of their energetic secondaries, upon nuclei of nitrogen and oxygen in the atmosphere. A good bit of evidence regarding the production processes is now available.

It was believed, in the early days of neutron work,[4,7] that most of the neutrons were produced as photo-neutrons by the numerous gamma-rays usually associated with the soft component of the cosmic radiation. There were three pieces of evidence which pointed toward this conclusion. First was the fact that neutrons increased with elevation at a rate[3] faster than the total radiation but at a rate just about the same as did the photons. Second was the fact that the photoneutron process was a known one, with reasonable cross sections so that the total number of neutrons could be accounted for without violating known numbers and distributions. Third were the direct experiments[10] in which actual coincidences were observed between neutrons counters and other kinds of counters arranged to detect the photon component. But a critical survey of the numbers of such coincidences led Korff[11] to the conclusion that this was probably not the only process possible. With the development of the new technique for observing cosmic-

[9] S. A. Korff and B. Hamermesh, Phys. Rev. **69**, 155 (1946). Korff, George, and Kerr, Phys. Rev. **73**, 1133 (1948).

[10] S. A. Korff, Phys. Rev. **57**, 555(A) (1940); Proc. Am. Phil. Soc. **84**, 589 (1941); S. A. Korff and E. T. Clarke, Phys. Rev. **61**, 422 (1942).

[11] S. A. Korff and B. Hamermesh, Phys. Rev. **71**, 842 (1947).

ray "stars" or nuclear disintegrations, it became apparent that there were very many more such disintegrations than had been at first recognized. A recalculation of the problem[11] led to the conclusion that a better fit was obtained if it was assumed that most of the neutrons came from such nuclear disintegrations and a smaller fraction from photo-neutrons. The belief today,[12] therefore, is that the bulk of the neutrons are produced in stars, and that while there are certainly some photo-neutrons, this is not the main source of them. Further, V. Tongiorgi Cocconi[13] showed that some neutrons are also associated with giant showers. This again provides another source, but not the main one.

Another experimental support of the view that many neutrons originate in stars was brought out in neutron coincidence experiments. Since stars are events in which two or more protons emerge from a nucleus, it is to be expected that neutrons will also be produced more than one at a time. Korff and Hamermesh,[14] therefore, set up two neutron counters in coincidence and found an appreciable counting rate of doubles at high elevations. Since the neutron is used up in producing a count, and so cannot produce two, any event in which both counters discharge simultaneously indicated two neutrons. The usual precautions were taken to eliminate background, chance coincidences and other spurious events.

Since the cosmic radiation shows a latitude effect, it is natural to seek one in the neutrons too. In the early experiments, detection sensitivity was not sufficient to reveal one, but in a series of flights in a B-29 at between 30,000 and 40,000 ft, J. A. Simpson[15] found such an effect. The latitude effect has also been observed in balloon flights to higher altitudes. Since the neutrons are secondary particles, the latitude effect in the neutrons is also the latitude effect in whatever process produces them. Further, since the latitude variation in neutrons is greater than that of the total intensity, it is clear that the neutrons are especially dependent upon the latitude-sensitive component of the total radiation. Since the number of cosmic-ray particles incident upon the outer parts of the atmosphere in unit time and per unit area increases rapidly as we go farther north, and since in this same latitude interval the average energy per particle decreases somewhat more slowly, it appears that the neutron production process does not depend so critically on the energy in this energy-interval. Thus, for example, if we consider a two-billion-volt primary particle, it will produce, on the average, a certain number of neutron secondaries. But a four-billion-volt primary will apparently not produce twice as many. The details of this situation, and the exact numbers involved, are under study at the present time. Further information is required, and it is to be hoped that experiments now in progress or to be made in the next few years will throw a lot of light on this rather complex situation.

At the present time it looks as though what was happening could be summarized thus: A primary is incident upon a nucleus in the upper atmosphere. From this nucleus it produces a number of high energy secondaries. If the primary energy is $10^{10}$ ev, the average energy of the secondary nucleons will be a few times $10^8$ ev. But now these same secondary nucleons can, by further nuclear collisions, produce still more nuclear particles, i.e., protons and neutrons. Each one, colliding with a nitrogen nucleus, has enough energy to disrupt it rather completely; and even if it is not completely disrupted, several nucleons will be ejected. It is these secondary disruptions which we see in emulsions as the "stars." The particles emerging in these stars are of energies of a few Mev, and so they leave heavy tracks in case they are protons. If they are neutrons, which they will be at least as often, they will not be visible in the emulsions or detectible in cloud chambers, but will actuate neutron counters. Here then we have the source of the bulk of our neutrons and an outline of the operation of the "nuclear cascade."

The bulk of the work done in the United States on this problem has been under a program sponsored by the Office of Naval Research and the Atomic Energy Commission. These two organizations have made intelligent provision for the stimulation of a great deal of excellent scientific work.

[12] S. A. Korff, Revs. Modern Phys. **20**, 327 (1948), and **21**, 75 (1949).
[13] V. T. Cocconi, Phys. Rev. **74**, 226 (1948); Cocconi, Cocconi, and Greisen, Phys. Rev. **74**, 1867 (1948).
[14] S. A. Korff and B. Hamermesh, Phys. Rev. **70**, 429 (1946).
[15] J. A. Simpson, Phys. Rev. **74**, 1214(A) (1948).

# On Illustrating the Regularity of Alpha-Radioactivity*

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The regularity of alpha-radioactivity is usually illustrated by the Geiger-Nuttall Law. Guided by the quantum-mechanical solution of the problem of alpha-particle emission, a way of plotting these data is shown, which yields a more satisfactory straight line than the Geiger-Nuttall plot.

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THE regularity observed in alpha-radioactivity between the half-life of an alpha-emitter and the range or energy of the alpha-particle emitted has been the subject of many investigations over the past forty years. In 1911, Geiger and Nuttall[1] found empirically that if the logarithm of the half-life of each element was plotted against the range of the alpha-particle emitted, an approximately straight line was found for each radioactive series. As recently as 1950, Perlman, Ghiorso, and Seaborg[2] reported at length on the systematics of alpha-radioactivity. It is the purpose of this paper to present a way of plotting data of this sort, which yields a more satisfactory straight line than the usual plot of the Geiger-Nuttall law.

TABLE I.† Well-authenticated alpha-emitters, with energies and half-lives. These isotopes do not emit gamma rays, do not suffer $K$-capture, and lose negligible amounts of energy (if at all) in the form of beta-radiation.

| Element | Energy-Mev | Half-life |
|---|---|---|
| Polonium-211 (AcC′) | 7.434 | 0.005 sec |
| Polonium-212 (ThC′) | 8.776 | $2.2\times10^{-7}$ sec[a] |
| Polonium-213 | 8.336 | $4.2\times10^{-6}$ sec |
| Polonium-214 (RaC′) | 7.680 | $1.5\times10^{-4}$ sec |
| Polonium-215 (AcA) | 7.365 | $1.83\times10^{-3}$ sec |
| Polonium-216 (ThA) | 6.774 | 0.158 sec |
| Polonium-218 (RaA) | 5.998 | 3.05 min |
| Astatine-215 | 8.00 | $10^{-4}$ sec |
| Astatine-216 | 7.79 | $3\times10^{-4}$ sec |
| Astatine-217 | 7.02 | 0.018 sec |
| Radon-217 | 7.74 | $10^{-3}$ sec |
| Radon-218 | 7.12 | 0.019 sec |
| Radon-219 (An) | 6.824 | 3.92 sec |
| Radon-220 (Tn) | 6.282 | 54.5 sec |
| Radon-222 | 5.486 | 3.825 days |
| Francium-219 | 7.30 | 0.02 sec |
| Francium-220 | 6.69 | 27.5 sec |
| Francium-221 | 6.30 | 4.8 min |
| Radium-221 | 6.71 | 31 sec |
| Radium-222 | 6.51 | 38 sec |
| Radium-224 (ThX) | 5.681 | 3.64 days |
| Actinium-225 | 5.80 | 10.0 days |
| Thorium-226 | 6.30 | 30.9 min |
| Thorium-229 | 4.85 | 7000 years |
| Thorium-232 | 3.98 | $1.39\times10^{10}$ years |
| Uranium-230 | 5.85 | 20.8 days |
| Uranium-232 | 5.31 | 70 years |
| Uranium-234 (U-II) | 4.763 | $2.35\times10^{5}$ years |
| Uranium-238 | 4.180 | $4.51\times10^{9}$ years |
| Neptunium-237 | 4.77 | $2.20\times10^{6}$ years |
| Plutonium-241 | 5.0 | 10 years |
| Plutonium-240 | 5.1 | 6000 years |
| Plutonium-238 | 5.51 | 92 years |
| Plutonium-236 | 5.75 | 2.7 years |
| Curium-240 | 6.26 | 26.8 days |
| Curium-242 | 6.08 | 150 days |
| Californium-244 | 7.1[b] | 45 min[b] |

† Data taken from reference 8.
[a] F. W. Van Name, Jr., Phys. Rev. 75, 100 (1949).
[b] Thompson, Street, Jr., Ghiorso, and Seaborg, Phys. Rev. 78, 298 (1950).

The Geiger-Nuttall law is usually illustrated by plotting the logarithm of the half-life of the alpha-emitter against the energy of the alpha-particle emitted. The curve which results when the data of Table I are plotted in this way is shown in Fig. 1. It is seen that the curve departs greatly from linearity and that furthermore, there is considerable scatter of the points about the curve. The regularity exhibited by this curve is hardly convincing.

The quantum-mechanical theory of alpha-emission was first treated, in 1928, by Gamow[3] and by Condon and Gurney,[4] and many more refined theories have been devised since then. Part B of Bethe's[5] review article lists various references through 1937. In all of the theories, the problem is to compute the probability that an alpha-particle inside of the nucleus will "leak through" the nuclear potential barrier and suddenly appear outside of the nucleus. The proba-

* This is the substance of a paper presented to the annual meeting of the Pennsylvania Conference of College Physics Teachers at Haverford College, Haverford, Pennsylvania, on October 21, 1950.

[1] H. Geiger and J. M. Nuttall, Phil. Mag. 22, 613 (1911).
[2] Perlman, Ghiorso, and Seaborg, Phys. Rev. 77, 26 (1950).
[3] G. Gamow, Z. Physik 51, 204 (1928).
[4] E. U. Condon and R. W. Gurney, Nature 122, 439 (1928).
[5] H. A. Bethe, Revs. Modern Phys. 9, 69 (1937).

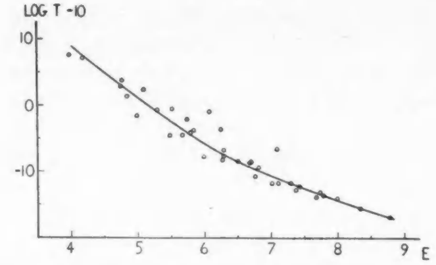


FIG. 1. Geiger-Nuttall curve; the logarithm of the half-life of each alpha-emitter ($\sec \times 10^{-10}$) is plotted *vs* the energy of the corresponding alpha-particle emitted (Mev).

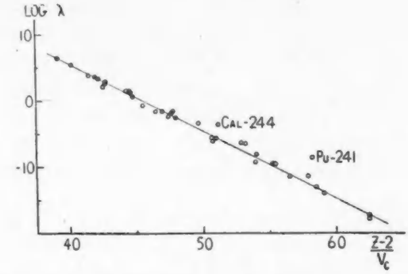


FIG. 2. Curve based on Gamow's quantum-mechanical theory of alpha-particle emission; the logarithm of the disintegration constant (per second) is plotted *vs* the quantity $[(Z-2)/v_c]\times 10^9$, where $Z$ is the atomic number of the emitter and $v_c$ is the velocity of the alpha-particle in cm/sec.

bility per second of this occurring then gives the fraction of such nuclei which will disintegrate per second, which is the disintegration constant, $\lambda$.

Perhaps the most rigorous calculation of the relations between the various parameters involved in alpha-radioactivity was by Preston.[6] Unfortunately, Preston's theory is too complicated to be represented graphically for the general case, where there may be accompanying emission of gamma-radiation and changes in the nuclear spin. However, for the case of nuclei which do not emit gamma-rays, Preston's theory is equivalent to a simpler theory developed by Gamow.[7] For simplicity, this paper will consider only elements which do not emit gamma-radiation, and Gamow's theory will be used.

The expression given by Gamow for the disintegration constant of an emitter of alpha-particles is as follows:

$$\ln\lambda = \ln(h/4mr_0^2) - \left(\frac{8\pi^2 e^2}{h}\right)\left(\frac{Z-2}{v_c}\right) + \frac{16\pi e}{h}[mr_0(Z-2)]^{\frac{1}{2}}, \quad (1)$$

where $\lambda = 0.693/T$ = disintegration constant ($\sec^{-1}$), $Z$ = atomic number of emitter, $v_c$ = velocity of alpha-particle relative to the nucleus (cm/sec), $r_0$ = nuclear radius (width of nuclear potential barrier) (cm), and $m$ = mass of alpha-particle (g). If values of the physical constants are substituted into Eq. (1) and common logarithms used, we find:

$$\log\lambda = -3.61 - 2\log r_0 - 1.192\times 10^9 (Z-2)/v_c + 4.23\times 10^6 [r_0(Z-2)]^{\frac{1}{2}}. \quad (2)$$

In Eq. (2) the velocity of the alpha-particle relative to the nucleus is computed from the observed value of the velocity by using the equation:

$$v_c = v_{\mathrm{obs}}[1 + 4/(A-4)]. \quad (3)$$

Let us now consider the various terms in Eq. (2). The term involving $\log r_0$ can be treated as essentially constant, since it will probably not change greatly from one element to another. Similarly, we may be able to neglect the variation of the last term, provided that the values of $(Z-2)$ and $r_0$ do not vary too much among the radioactive elements. This suggests that if we plot $\log\lambda$ *versus* $(Z-2)/v_c$, a straight line should result.

The data to be plotted to illustrate the foregoing discussion were taken from the table of isotopes in Bitter's[8] recent textbook on nuclear physics, except where noted. The criteria for inclusion in Table I were that the isotope identification be certain, that no accompanying gamma-radiation be emitted, that $K$-electron capture not take place, and that if any beta-particle is emitted by the nucleus, its energy be negligible. It would be expected that alpha-emitters which met these requirements would obey the simple Gamow formula. Thirty-seven isotopes of this type were found.

[6] M. A. Preston, Phys. Rev. **69**, 535 (1946); **71**, 865 (1947).

[7] G. Gamow, *Atomic Nuclei and Radioactivity* (Oxford University Press, New York, 1931), p. 51.

[8] Francis Bitter, *Nuclear Physics* (Addison-Wesley Press, Cambridge, Massachusetts, 1950).

Figure 2 shows the plot of the data in Table I using the Gamow formula. The resulting curve is seen to be quite a good straight line, except for Plutonium-241 and Californium-244. However, the energies of these elements are known only to two significant figures, and it is seen from the expanded scale of the abscissa that an error of several percent in the velocity of the alpha-particle would produce discrepancies of the size found. On the whole, the Gamow formula seems to be well verified, and this diagram seems to be clearly better than the Geiger-Nuttall plot for demonstrating the regularity of alpha-radio-activity.

# Demonstration Experiments in Electromagnetic Induction

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(Received June 16, 1950)

Several standard demonstration experiments in electromagnetic induction, as performed with newly designed or modified apparatus are described. One of these, Arago's disk, employs a rotating magnetic field so that torque measurements can be made on the (almost stationary) disk. A series of auxiliary measurements shows that the torque $T$ can be represented in terms of the flux $\phi$ by the relation $T=1.55\phi^2$, when $T$ is in g wt×cm, and $\phi$ is in kilogauss.

THE experiments outlined in this paper are, for the most part, modifications of those described in elementary textbooks. For example, to extend the experimental range of Arago's disk experiment on electromagnetic induction, an apparatus (illustrated in Fig. 1) was constructed with an electromagnet $M$ replacing the permanent magnet used in Arago's experiment. The electromagnet was constructed with four pole pieces attached to an iron plate $P$, which was mounted on the end of a shaft connected through a gear assembly $G$ to an electric motor. Low resistance coils were wound around the pole pieces with the direction of the winding alternating. These were joined in series and connected through a brush $B$ and slip ring $R$ to an external circuit consisting of a 6-volt battery in series with a variable rheostat. The frame of the motor was used as the other connection between the battery and the electromagnet. The speed of the motor driving the unit was controlled by the usual methods employed for electric motors.

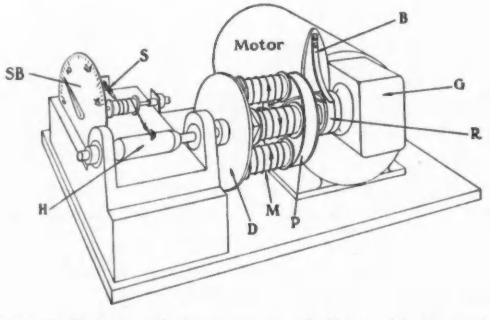


FIG. 1. Rotating electromagnet and disk to illustrate the action of a rotating magnetic field on a conductor.

A circular copper disk $D$, 3.6 in. in diameter and $\frac{5}{64}$ in. thick, mounted on a shaft is supported with its face opposite the pole pieces of the electromagnet. The spring balance $SB$ is used to measure the torque exerted on the disk by the rotating magnetic field. The balance may be disconnected at $H$ for demonstrations in which the free rotation of the disk is desirable.

Two variations in experimental procedure were employed with this apparatus. In the first case, the electromagnet current was kept constant and the speed of rotation was varied. For this work the torque, as registered by the spring balance, was proportional to the speed of rotation. In the second test, a constant speed of rotation was maintained and the current in the electromagnet was varied. The increase in torque with current was very marked and, with a rotational speed of 1060 rpm, it varied from 3.9 to 60 g wt×cm with the current increasing from 0.5 to 1.8 amp.

An auxiliary set of measurements was made, by means of a fluxmeter connected to search coils circumscribing the poles of the electromagnet, to determine the relation between the current in the electromagnet and the average magnetic flux per pole. A series of values is tabulated in Table I.

A logarithmic graph shown in Fig. 2 was plotted for these values. The deviations of the points from a straight line are within the range of the deviations involved in measuring the torque $T$. Equation (1), derived from the graph, represents the relationship between $T$ and $\varphi$ with a fair degree of accuracy:

$$T = 1.55\varphi^2. \qquad (1)$$

To illustrate the action of laminations in electrical apparatus, a piece of equipment illustrated in Fig. 3 was constructed. The essential component is a 1-in. cube constructed of thin square copper plates held together with radio cement. This was mounted in a frame with a plastic base and a top bar, provided with a hook, and clamped to the base by means of two small brass bolts.

TABLE I. Representative values of the current and magnetic flux of the electromagnet, the spring balance dial readings, and torque for a speed of rotation of 1060 rpm.

| Current in the electromagnet (amp) | Magnetic flux per pole $\phi$ (kilogauss) | Average spring balance dial readings | Torque $T$ (g wt × cm) |
|---|---|---|---|
| 0.5 | 1.58 | 2.3 | 3.9 |
| 0.6 | 1.98 | 3.0 | 5.4 |
| 0.8 | 2.66 | 5.8 | 11.2 |
| 1.0 | 3.45 | 8.5 | 17.8 |
| 1.25 | 4.35 | 12.4 | 27.8 |
| 1.50 | 5.25 | 18.0 | 42.2 |
| 1.80 | 6.20 | 24.5 | 60.0 |

To carry out the demonstration, the equipment is suspended by means of a thread with the cube between the poles of a powerful electromagnet supplying a magnetic field in a horizontal direction. With the cube in the position indicated in Fig. 3, the plane of the laminations is parallel to the direction of the field; and, consequently, there is high electrical resistance in the direction in which the induced emf acts. This reduces the eddy currents to negligible values, and, hence, the damping effect is small. For the second part of the demonstration, the cube is remounted in the frame with the laminations perpendicular to the base, and the experiment is repeated. This time the damping effect as the cube spins in the field is very marked, since the plane of the laminations is parallel to the direction in which the induced emf acts.

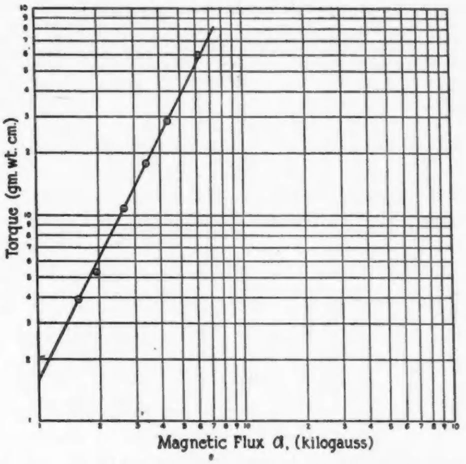


FIG. 2. Logarithmic graph of torque on disk *vs* magnetic flux in rotating magnet.

Spectacular and interesting demonstrations are secured by means of the jumping ring experiment.[1] The standard Cenco equipment provided for this experiment consists of an electromagnet with a laminated iron core 12 in. long and 1 in. in diameter, designed for use in a 60-cycle ac circuit, and an assortment of cylindrical aluminum and copper rings. To carry out the test, the core is raised up and a ring is put

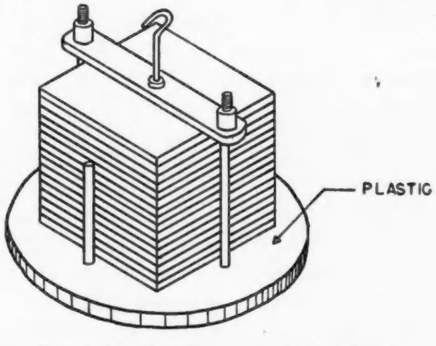


FIG. 3. Laminated copper cube and frame.

[1] R. M. Sutton, *Demonstration Experiments in Physics* (McGraw-Hill Book Company, Inc., New York, 1938), pp. 349–350.

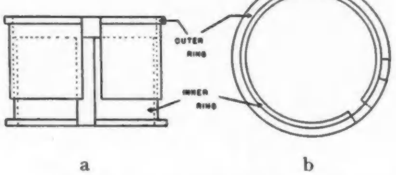


FIG. 4. Concentric cylindrical rings.

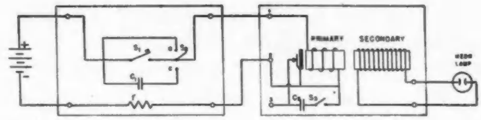


FIG. 5. Circuit designed for experimental work with small induction coils.

around it. A lively jump is then secured when the electromagnet is connected into the circuit. If a split ring is used, no effect is observed.

A double-ring device (illustrated in Fig. 4) was constructed to act either as a closed or split ring. It consists of two concentric copper rings with an outside diameter of $2\frac{1}{4}$ in., a thickness of $\frac{1}{16}$ in. for each, and a height of 1 in. These were carefully machined in a lathe to secure a close fit, and slots were cut in each ring. With the apparatus as shown in Fig. 4a, it acts as a split ring and when it is used with an electromagnet in an ac current, no action is secured. By simply rotating the rings so as to close the gap, as in Fig 4b, the ring shoots up.

Another variation often employed with this experiment consists of cooling the ring in liquid air before carrying out the test. The increased conductivity augments the jump. One of the most important points in the experiment, which apparently has not been stressed, is the position of the iron core. Tests with the standard Cenco magnet showed that the best results were secured with the extension of the core between 9 and 10 in. above the top of the solenoid. Using a light aluminum ring, with the core in this position, jumps of over 7 ft were secured at room temperature, and jumps of 14 ft with the ring precooled in liquid air. This brought the ring close to the ceiling above the lecture table, so no attempt was made to increase the jump by redesigning the electromagnet. The impedance of the electromagnet is small with the high extension of the core; and, therefore, it should not be left connected in the circuit for a longer time than required to carry out the demonstration.

The principles of electromagnetic induction are exemplified in an excellent manner by the action of the induction coil. A circuit (illustrated in Fig. 5) which was designed for use with elementary classes for the study of the induction coil has proved very useful with large classes for which cathode-ray oscilloscopes were unavailable.

The induction coil employed with this circuit is a small unit with a standard primary and a special secondary coil consisting of 9800 turns of No. 36 enamelled copper magnet wire wound on an ebonite spool 3 in. long. This was supported on two brass rods so as to allow variation in the distance of separation and the inclination of the coils relative to each other. These rods also acted as bus bars for the secondary circuit which was connected to a 2-watt neon lamp with a sparking potential of 72 volts dc. A small protective resistance $r$ was included in the primary circuit which was operated by a 6-volt storage battery.

The connections, as shown in Fig. 5, are used to determine the respective couplings of the coils required to secure a discharge in the neon lamp on the break and make of the primary circuit by means of the press key $S_1$. With switch $S_2$ thrown to $c$, a 0.25-$\mu$f condenser $C_1$ is connected in parallel with the switch $S_1$, and tests can be made to determine the action of the condenser.

For work with the automatic make and break, the switch $S_2$ is thrown to $a$, and the connecting wire on post 2 is changed over to post 3. The 0.25-$\mu$f condenser $C_2$ is controlled by the switch $S_3$. Using the automatic make and break with the condenser $C_2$ in parallel, the secondary emf was sufficient to give a unidirectional discharge in the neon lamp for couplings of the coil equivalent to a separation of 11 cm when parallel to each other or 8 cm with an inclination of 60° in the secondary. For a discharge in both directions, a much closer coupling corresponding to separation of 4 cm was required. This equipment can be used not only to illustrate the essential features required for an efficient induction coil but also to illustrate the variation in the secondary emf with the coupling between the coils.

# Dispersion and Resolving Power of Prism Spectrometers


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(Received October 9, 1950)


Brief derivations of dispersion and resolving power give their expression in terms of the directly measurable angle of the prism and its angle of deviation at different wavelengths. Elimination of the length of the prism base seems to improve student understanding of the quantities. Students' experimental results are given to show the precision to be expected.


FOR some time in our junior-senior laboratory course we have measured the dispersion and resolving power of a prism spectrometer using a flint glass prism with an index of refraction of 1.65. To do this the student must use relations involving the length of the base of the prism used. In actual practice, since the light beam from the collimator does not usually cover the whole face of the prism, the student must determine the maximum distance the beam traveled through the prism. This unfortunate situation is apparently the result of the original Rayleigh formulation, which is repeated in laboratory manuals and textbooks of physical optics. The expression of dispersion and resolving power in terms of the base of the prism is partially modified in only one book.[1] The writer hopes that a presentation more in line with experiment will be useful in helping students to gain familiarity with the concepts of dispersion and resolving power.

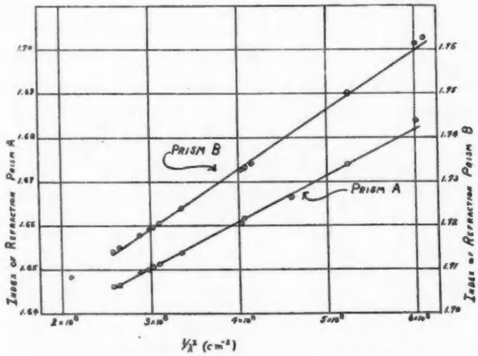


Fig. 1. Index of refraction *vs* reciprocal of the square of the wavelength ($cm^{-2}$).

[1] R. A. Houston, *Treatise on Light* (Longmans, Green and Company, New York, 1938), fifth edition, p. 238.

## DISPERSION

The index of refraction of a given prism for different wavelengths in a spectrum can be found in the usual way by the relation

$$\mu = \sin\tfrac{1}{2}(A+\delta)/\sin\tfrac{1}{2}A, \qquad (1)$$

where $\mu$ is the index of refraction, $A$ is the refracting angle of the prism, $\delta$ is the angle of minimum deviation for a given wavelength $\lambda$. By correlating indices of refraction with the corresponding wavelengths one can plot a dispersion curve and thus find the slope $d\mu/d\lambda$ of this curve as a function of $\lambda$. We have found it more convenient, since only the visible spectrum is used, to assume the cauchy dispersion formula

$$\mu = \alpha + \beta/\lambda^2, \qquad (2)$$

and obtain $\beta$ from the linear plot of $\mu$ *vs* $1/\lambda^2$. (See Fig. 1.) Then we have

$$d\mu/d\lambda = -2\beta/\lambda^3. \qquad (3)$$

Depending on one's preference, the dispersion $D$ of the prism can be defined as either $d\delta/d\lambda$ or $di'/d\lambda$, the angles $\delta$ and $i'$ being indicated in Fig. 2. Taking as our definition

$$D = d\delta/d\lambda = (d\delta/d\mu)(d\mu/d\lambda), \qquad (4)$$

we need to know only $\delta$ as a function of $\mu$ to obtain the dispersion. This function is obtained

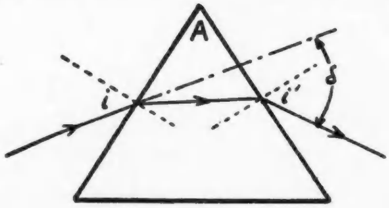


Fig. 2. Refraction through a prism: Notation for angles used in discussion.

TABLE I. Dispersion of prisms.

| Wavelength (A) | Prism A Index of refraction | Prism A Dispersion | Prism B Index of refraction | Prism B Dispersion |
|---|---|---|---|---|
| 4077 | 1.684 | 5840 | 1.762 | 8300 |
| 4343 | 1.674 | 4760 | 1.750 | 6780 |
| 4393 | 1.667 | 4560 | | |
| 4960 | 1.662 | 3150 | | |
| 4981 | 1.661 | 3120 | 1.733 | 4370 |
| 5461 | 1.654 | 2340 | 1.724 | 3270 |
| 5685 | 1.651 | 2060 | 1.721 | 2880 |
| 5770 | 1.651 | 1970 | 1.620 | 2760 |
| 5791 | 1.650 | 1950 | 1.719 | 2720 |
| 5893 | 1.649 | 1850 | 1.718 | 2570 |
| 6158 | 1.646 | 1610 | 1.715 | 2240 |
| 6234 | 1.645 | 1550 | 1.714 | 2160 |
| 6907 | | | 1.708 | 1570 |

TABLE II. Resolving power of prisms.

| Wavelength (A) | Prism A R (exper.) | Prism A R (calc.) | Prism B R (exper.) | Prism B R (calc.) |
|---|---|---|---|---|
| 4339.2<br>4347.5 | 585 | 525 | | |
| 4979.0<br>4983.2 | 910 | 1185 | 1340 | 1185 |
| 5682.8<br>5688.3 | | | 1000 | 1030 |
| 5769.6<br>5790.7 | 255 | 275 | 310 | 275 |
| 5890.0<br>5895.9 | | | 975 | 1000 |
| 6154.4<br>6160.8 | | | 1035 | 965 |

from Eq. (1), which gives, on differentiation,

$$d\mu/d\delta=\tfrac{1}{2}\cos\tfrac{1}{2}(A+\delta)/\sin\tfrac{1}{2}A. \qquad (5)$$

Substitution of Eqs. (5) and (3) in Eq. (4) gives

$$D=-\frac{2\sin\frac{1}{2}A}{\cos\frac{1}{2}(A+\delta)}\cdot\frac{2\beta}{\lambda^3}. \qquad (6)$$

Thus, once the indices of refraction are measured by the student, sufficient information is available for him to compute the dispersion as a function of wavelength for his prism. Typical student data and results are shown in Table I.

## RESOLVING POWER

In the case of resolving power we use the Rayleigh criterion; that is, the two lines of a doublet are considered resolved if the central maximum in the diffraction pattern of one line lies at the first minimum in the diffraction pattern of the second line. In the case of a prism two lines will be resolved if the angular separation $d\delta$ of the two lines is the same as the angular half-width of the Rayleigh criterion. To determine when this is so, one puts a variable width slit directly in front of the spectrometer telescope and reduces the slit width until a given pair of lines (any doublet) merge into one. If the slit width at this point is $b$, the angle of diffraction will be $\lambda/b$ and equal to the angular separation $d\delta$ produced by the prism; or, the angular resolving power will be

$$R_a=b/\lambda=1/d\delta,$$

or

$$bd\lambda/\lambda=d\lambda/d\delta,$$

from which we see that the theoretical definition of resolving power as $\lambda/d\lambda$ can be expressed as

$$R=\lambda/d\lambda=bd\delta/d\lambda.$$

Further, since $d\delta/d\lambda$ is the dispersion, we have the familiar relation that

$$R=bD.$$

By choosing various doublets in the sodium and mercury spectra, the student can easily compare calculated and measured resolving powers and see how seriously the optical system of the spectrometer affects the theoretical result. Table II shows student data and results of such a test.

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*It was not the philosopher Kant, but the mathematician Gauss who first discovered that the three angles of a triangle might not add up to two right angles. It was the physicist Einstein who gave life to the principle that our theories should not be concerned with things that are unobservable.*
—C. G. DARWIN, The New Conceptions of Matter, 1931.

# Compressible Flows*

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(Received August 21, 1950)

A brief exposition (descriptive rather than quantative) is given of some of the common phenomena of compressible flow. The aim of the paper is to supply background material to the physicist who may wish to know a little more about the problems and techniques involved in the mechanics of compressible fluids.

A graphical representation of a point source moving through a fluid at various velocities from subsonic to supersonic is given and described. There is also a description with sketches of the principal optical methods of wind-tunnel testing for compressible flows. A number of actual test photographs is included which show the type of picture obtained by the various test methods and which also show the formation of shock and Mach waves.

THE rather strange combination of military necessity and Sir Isaac Newton has contributed much to the advancement of modern science. The pressure of national emergencies and wars has accelerated scientific research and contributed materially to both new discoveries and new developments. In the background seemingly always ready either with part of the answer or as the one who first attacked the problem is Sir Isaac.

The study of compressible flow is no exception, for it was Newton who made the first theoretical determination of the velocity of sound. His calculations were in error, as he knew from the results of ballistic experiments. Subsequently, Laplace made the correct determination, and it was seen then where Newton's error lay. Newton had assumed isothermal conditions, whereas it is now known that adiabatic conditions are much closer to the true situation.

The greatest mass study of compressibility was initiated in the field of projectiles and has since been augmented by investigations of the high speed airplane and the various types of missiles. Interest in the phenomena of compressible flows is no longer limited to the theoretical physicist and applied mathematician. Today these phenomena are of equal, if not greater, importance to the aeronautical and mechanical engineers engaged in airplane and power plant design.

Up until a few years prior to World War II, it was assumed in practical aerodynamics that air could be considered to be incompressible, an assumption which, while *never* true at any velocity, was nevertheless reasonable enough when the magnitude of the errors thus introduced was compared to the magnitudes of other inherent inaccuracies. The error which the assumption of incompressibility of the air introduces can be given, approximately, as one-half the square of the Mach number. The Mach number is defined as the ratio of the velocity of a body in a medium to the velocity of sound in the same medium; it is called the Mach number for Ernst Mach, the Austrian physicist, who first used it. (The velocity $C$ of sound in air in feet per second can be quickly calculated from this simple formula: $C=49\times T^{\frac{1}{2}}$, where $T$ is expressed in degrees Rankine.)

When the term "error" is used in connection with the airplane or other bodies which move through any medium at high velocities, it means that the computed forces—drag, lift, etc., in the case of the airplane—may be in considerable error if the effect of compressibility of the air is neglected. For example, an airplane traveling at

* This paper is deliberately descriptive. The tendency in much of present-day physics is to reduce everything to mathematical equations so that it is perfectly possible to train human computing machines who think they are scientists. There is no reason why some of these highly involved (mathematical) topics cannot be "discussed" without losing sight of either the elegance or marvels of mathematical manipulation and formulation. The fact that such things as circulation (in lift theory) are largely mathematical concepts which cannot be completely explained physically means only that to date we have failed to find the satisfactory explanation, but under no circumstances imples that one cannot be found. Sometimes it is the "lazy" thing to jump into mathematics and forget the physics. The purpose of this paper is to try to discuss in a few words what investigators are doing in the world of compressible flows, the problems that face them, and the techniques being used to overcome or avoid these problems.

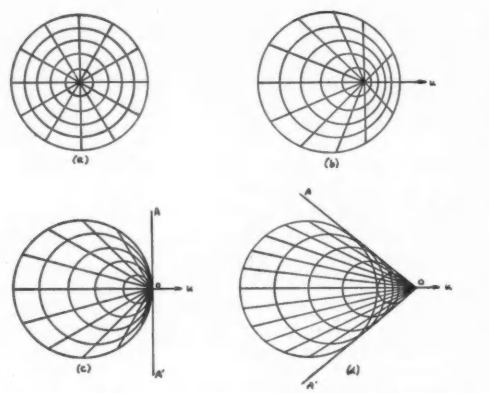

FIG. 1. A point source moving through a fluid. (a) The point is at rest in the fluid. (b) The point is moving at a velocity about half that of sound. (c) The point is moving at the velocity of sound. (d) The point is moving at a velocity about one and a half times as great as that of sound.

250 mph at sea level under standard atmospheric conditions would have a Mach number of about 0.33. If the effect of compressibility is neglected, the forces acting on this airplane might be in error by as much as 5 or 6 percent. In the present air age, only the "old-timers" travel at such slow rates. Therefore, any attempt to ignore the compressibility effect could be fatal.

## MACH ANGLE AND MACH CONE

Figure 1 gives a quick visual picture of what is happening in the four cases where a point source moves through a fluid at different velocities. If Fig. 1 is considered to be a plane representation of a three-dimensional figure then Fig. 1(a) represents the case in which concentric spherical surfaces are achieved by the pressure effect in equal time intervals. In Fig. 1(b) the centers of these surfaces are now moving relative to the medium, but the velocity is less than sonic. In Fig. 1(c) the surfaces are now tangent to each other at a common point, and the so-called Mach cone appears for the first time, in this case as a straight line, or a plane, as shown by the line *AOA'*. The Mach angle, which is defined as half the vertex angle of the Mach cone, is obviously equal to 90° in this case. Figure 1(d) shows what happens when the Mach number is greater than unity or when the velocity of sound has been exceeded. The line *AOA'* is now "bent" into the form shown; for the three-dimensional case a cone results instead of a triangle. Numerous investigators, among them Theodore von Kármán, have called the region outside the Mach cone the "zone of silence" and the region inside the cone the "zone of action."

What this terminology means is that the point source herein described can influence only those points which are within the zone of action, or, in the language of the aerodynamicist, those points which are downstream of the point source. For subsonic velocities the concentration of action is dispersed, as can be seen by the rays which emanate from the point source in Figs. 1(a) and 1(b). In Figs. 1(c) and 1(d) there is a grouping of this concentration within the region of the Mach cone. Therefore, the total effect tends to be localized within a relatively small area. As the Mach number is increased the Mach angle is decreased, and the volume affected is further reduced thereby causing a still greater effect. It might be well to point out here that the shock wave, about which so much is said, and the Mach wave (the Mach cone is actually a wave front) are not the same except under very special conditions.

A shock wave may be defined as some particular small region in space, taken with reference to a moving body, which is capable of causing a violent change in the distribution of pressure, velocity, temperature, and density of the fluid which passes through it. In general, the Mach and shock waves are quite different, with the intensity of the former being considerably less than that of the latter. It follows then that the Mach angle is always less than the shock angle. The only exception to this rule is the "weak" shock wave—one caused by a small disturbance or a so-called infinitesimally small shock—which tends to approach the Mach wave in intensity.

Because of the importance of Mach number and because it actually records velocity and indirectly gives information concerning the intensity of a shock wave, a device known as a *Machmeter* has been introduced into modern high speed aircraft. The Machmeter, which is a specially designed airspeed indicator calibrated to read Mach number directly, tells the pilot at a glance his velocity relative to that of sound.

To the theoretical physicist or aerodynamicist

there are three regions of interest in the problem of compressible flows. These are: the subsonic, transonic, and supersonic. Although the Mach angle and Mach cone do not specifically appear prior to the sonic range, shock waves do exist much earlier.

## SHOCK WAVE PHENOMENA

In the short calculation mentioned earlier, it was seen that a considerable error in the magnitude of the forces acting on an airplane could be introduced at relatively slow speeds. Once an airplane exceeds 400 mph, the visible shock effect may still be negligible but errors approaching 20 percent in the computation of such forces as the drag of the airplane may be introduced. However, at Mach numbers of 0.800 (approximately 600 mph) or above, the shock effect is already noticeable. From this point to a value of the Mach number slightly larger than unity, a body may be said to be in the transonic range. This is frequently the most difficult range in which to make studies, because it is neither truly subsonic nor truly supersonic but rather possesses characteristics of both types of motion. It might be compared to the transition range between laminar and turbulent flow, where it is also impossible to determine exactly what conditions prevail. In this transonic range the Mach number will vary for different points along the body, and the variation will be such that neither region is clearly defined.

The practical aerodynamicist is, therefore, faced with an extremely difficult problem. He must either design surfaces in such a way that the transonic region can be bridged without danger of sudden increases in pressure—and consequently lose in lift—or else he must somehow try to predict where the local shock wave will occur and make his design accordingly. The prediction of the location of the local shock wave is again a tricky problem. Mathematically the complexity is so great that analysis becomes uncertain. On the other hand, experimental data cannot always be considered reliable, since one cannot be certain that the conditions of an experiment will always be met satisfactorily in practice. The designer is usually forced to design for the worst possible condition and then keep his hopes high.

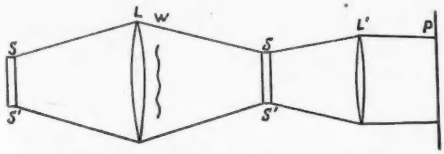

FIG. 2. Schematic diagram for a typical *schlieren* arrangement.

Since any attempt to discuss the mathematical methods used in the study of compressible flows would be out of place at this time, a brief discussion of the experimental methods will be introduced.

There are three optical methods in common use: the *schlieren* (streaks) method, the shadow method, and the interferometer method. Of the three, the most commonly used is the *schlieren* method, but the interferometer method is the best for quantitative analysis.

Figure 2 shows the schematic diagram for a typical *schlieren* arrangement. Here $SS'$ is a source of light, $L$ is a lens which causes the light to converge and form an image of the light source at $ss'$. The object to be examined is placed at $W$. In this case, it may be assumed to be the test section of a high speed wind tunnel. Every point on $W$ receives light from $SS'$ so that it is completely illuminated. Thus, light passing through $W$ goes on through a second converging lens $L'$ to a sensitized plate or screen located at $P$.

Now if an object such as a knife-edge is introduced into the field at the focal point $ss'$, there will be a dark space on the plate $P$ which corresponds to the knife-edge. Any bending of the light rays from $SS'$ will destroy the uniformly dark image of the knife-edge on $P$. Therefore, when the bent rays are again brought into focus on $P$ the illumination of that portion of the knife-edge will be greater or less depending on the direction in which the rays are bent. The change in density of the air in a wind tunnel, due to the appearance of a shock wave, will have the same effect as the introduction of a prism into the light path. The density change may be expressed as a simple density gradient. There are many variations and modifications of the *schlieren* method, but the above description gives the essential physical principles.

Figure 3 shows a typical shadowgraph arrangement. It is the simplest arrangement of

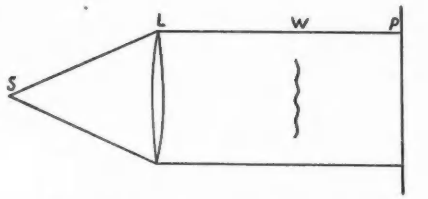


FIG. 3. Schematic diagram for a typical shadowgraph arrangement.

the three and gives remarkably good pictorial results, as Fig. 5 shows. Referring to Fig. 3; there is a source of light at $S$, a lens at $L$, the object to be studied at $W$, and a screen or plate at $P$. (The light is made parallel before passing through the object.) It is well known that the intensity of the light which falls on the plate will be a function of the density of the medium through which it passes. Mathematically the illumination turns out to be a function of the second derivative of the density or the first derivative of the density gradient.

The principle of the interferometer method is exactly that of the photoelastic methods, so much used in stress analysis. Since the setup is somewhat more complicated than in either of the above cases the details of construction will be omitted. It is possible to use the interferometer method for quantitative analysis, because the interference patterns can be studied and measured directly. The accuracy is about the same as that in photoelastic methods and gives, perhaps, to the theoretical investigator a hint and guide as to the direction in which he should move as well as a check on his calculations. Examples of all three types of experimental techniques are shown in the figures which follow.

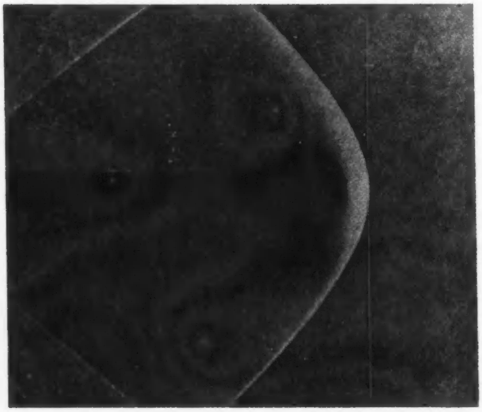

FIG. 4. Sphere in wind tunnel at Mach number 1.800. (Photograph is from Liepmann and Puckett, *Aerodynamics of a Compressible Fluid*, p. 96. See second reference in bibliography.)

Although the dimensions of the spheres in the photographs are not necessarily the same, the Mach numbers are almost identical. For Figs. 4 and 5 the Mach number is 1.80, and for Fig. 6 it is 1.81. In Fig. 4 the dark threaded arm shown at the left is merely the support for the sphere.

For qualitative purposes as well as for simplicity of the optical setup, the shadowgraph method certainly has a definite advantage, as can easily be seen in Fig. 4. The *schlieren* photograph of Fig. 5 offers nearly as good photographic results. However, the interference pattern as shown in Fig. 6 gives a much clearer picture, as well as the opportunity for quantitative examination, of what is happening to the medium surrounding the sphere.

The series of photographs shown in Fig. 7 gives a good picture of what happens to the air around a projectile as it travels through the air at velocities ranging from about 900 ft/sec to 2800 ft/sec. The formation of the shock waves and the turbulence of the wake show up very well. It can easily be seen how the "zone of action" is narrowed as the Mach number increases. These photographs of actual projectiles show how accurately the present theory has been able to describe the effect that high speed bodies have on the surrounding air. Designing to overcome the problems caused by shock waves is, of course, a tremendous additional problem.

## TEMPERATURE EFFECTS IN HIGH SPEED FLIGHT

Since kinetic energy is a function of velocity and since one manifestation of kinetic energy is temperature, it is not at all strange to find that high speed bodies possess some rather interesting heating effects. One of the most persistent problems of low velocity aerodynamics is the determination of the exact location of the point where the flow over the body changes from laminar to turbulent. This transition point is very important for many reasons, of which the problem of control is one of the foremost. Naturally enough, this

same problem exists in compressible flow problems, but it can be solved with reasonable ease because of the heating effect of the air as it passes the body. The method used in compressible flows is, of course, applicable in any case, whether of high or low velocity flight.

By measuring the temperature in the boundary layer itself, it is possible to determine where the change from laminar to turbulent flow occurs. The temperature in the turbulent region is higher than in the laminar region, and a simple formula which has the boundary layer temperature expressed as a function of the square of the Mach number gives the result very quickly. The only other factor which must be known is the temperature of the free stream. This method is better at the higher Mach numbers, because the constant which changes the formula from the laminar to the turbulent case varies in the second decimal place and is less than unity in both cases. Therefore, since this constant is multiplied by the square of the Mach number, the temperature difference will be small at low air speeds. However, at values close to a Mach number of unity and at all values above unity, the rise in temperature due to this skin friction effect is even greater than the empirical formula indicates.

For example, assuming an ambient air temperature of seventy degrees Fahrenheit and a Mach number of one-third, the difference between the temperatures in the laminar and turbulet portions of the flow is less than one degree Fahrenheit. At a Mach number of one, this may exceed ten degrees.

While the observers in laboratories may welcome this relatively simple method of locating the transition point in a flow, the pilots of high-speed aircraft find a rather different problem. In actual flight the interior of a cockpit may be turned into an inferno. Using a simple rule of thumb device which seems to hold for Mach numbers of 0.80 and above, if the number of hundreds of miles per hour if squared, the temperature in degrees Centigrade is found at once. Using this rule we find that the temperature inside an uninsulated cockpit might *increase* as much as 130°F at a Mach number of one (approximately 750 mph). This added to the outside air temperature in any place but a subzero zone would be fatal to a human being if he had to endure it for any period of time.

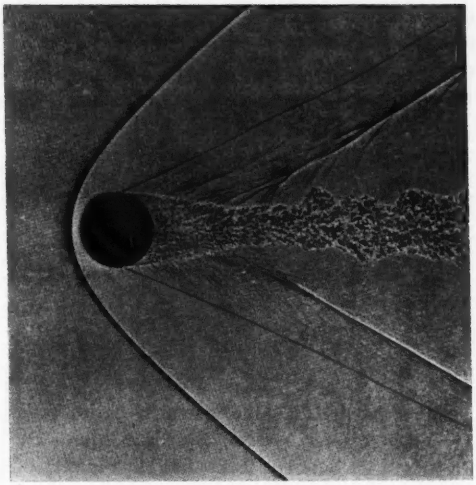

FIG. 5. Shadowgraph of a sphere in free flight at Mach number of 1.800. (Photograph from the Aberdeen Proving Grounds.)

Many modern high speed military aircraft are being equipped with refrigerating systems to help keep the pilot from being roasted alive. This, incidentally, was an effect which received almost no attention when the first high speed aircraft were designed. The seriousness is not only in pilot comfort and safety but also in danger to the airplane itself. Most airplanes are of aluminum sheet construction riveted with aluminum rivets. The strength properties of aluminum change

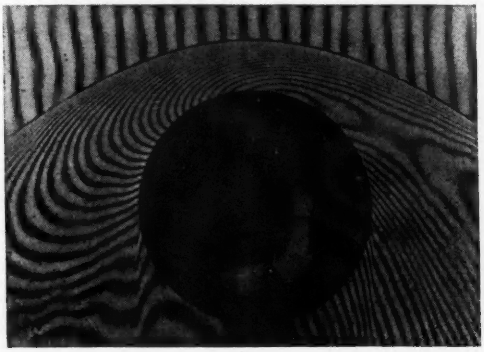

FIG. 6. Interferogram of the flow around a sphere at Mach number 1.81. (Photograph from N.A.C.A., T. N. 2000. See fifth reference in bibliography.)

greatly with temperature, falling off with an increase in temperature. Rivets heated up in this fashion have been known to "pop" in flight when no other undue stress was placed on them.

## CONCLUSIONS

One can gather from the foregoing that the problems involved in compressible flows are both

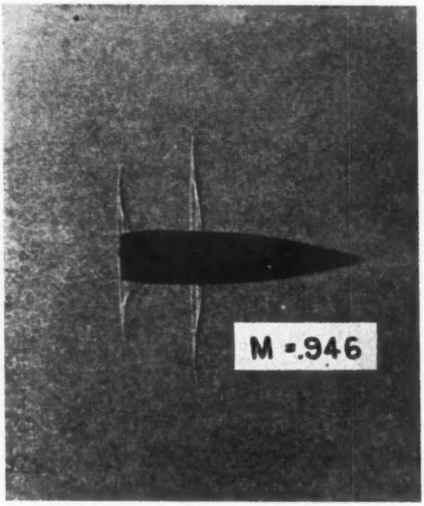


FIG. 7(a).


FIG. 7(b).

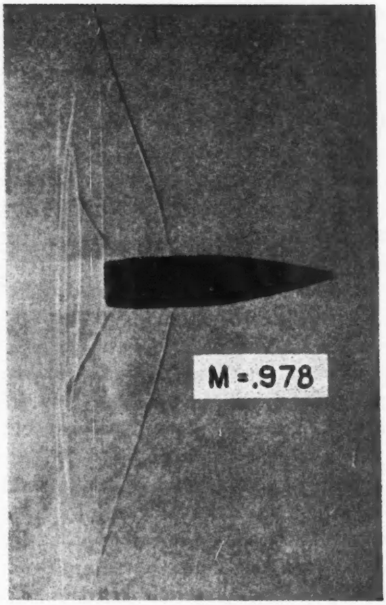


FIG. 7(c).

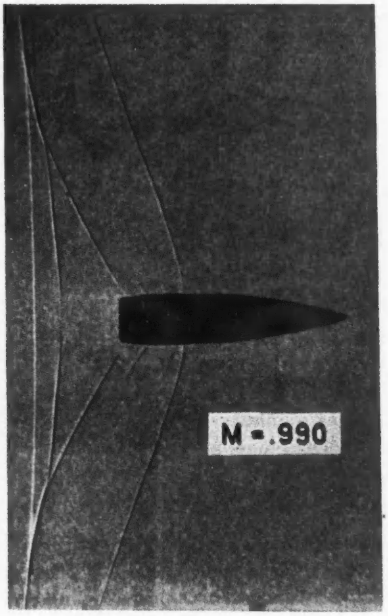


FIG. 7(d).

FIGS. 7(a) to 7(*f*). Shadow photographs of 155-mm projectile in free flight at Mach numbers *M* ranging from 0.946 to 2.479. (Photographs from the Aberdeen Proving Grounds.)


FIG. 7(e).

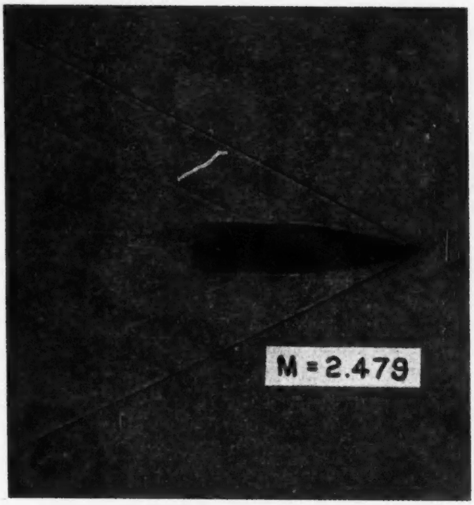


FIG. 7(f).

interesting and difficult. Theoretically and experimentally much remains to be done. The United States was, unfortunately, a very slow starter in this field; only one supersonic wind tunnel existed in this country until the last months of World War II, although there were several in operation in Germany with several dozen more in various stages of planning.

In the last five years the picture has changed quickly and radically. Numerous tunnels are either in operation or under construction in the United States, and the amount of theoretical investigation has increased by several thousand percent. However, next to nuclear research, high speed aerodynamic research on the experimental level is the most expensive type of work an organization can undertake. Therefore, nearly all of it is being done through government agencies, with relatively little done by private sources.

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James B. Kelley, Phys. Today **3**, No. 4, 20–24 (1950).

## Summer Meeting of the Association at Michigan State College

The 1951 summer meeting of the American Association of Physics Teachers will be held at Michigan State College, East Lansing, Michigan, on June 26–28. The American Society for Engineering Education is also to meet at Michigan State College over the period June 25–29. In some of the sessions the Association and the Society will meet jointly. Ample dormitory accommodations will be available.

Titles of 10-minute contributed papers to be presented at the meeting should be sent before May 7 to Dr. Robert F. Paton, Secretary AAPT, Department of Physics, University of Illinois, Urbana, Illinois.

# NOTES AND DISCUSSION

## An Inexpensive High Resistance Voltmeter

ORRIN H. SMITH
*DePauw University, Greencastle, Indiana*

PHYSICS teachers with limited departmental appropriations have frequently felt a need for a voltmeter of high resistance approaching the conditions of an electrostatic instrument. The modern multimeter and the vacuum tube voltmeter serve very well in this field for departments that can possess them.

The author has found a fairly simple and inexpensive device that is quite flexible and accurate and is semiportable and that can be made up easily in the simplest shop. It is assumed that the department already possesses a Leeds and Northrup type P galvanometer or its equivalent. The added materials consist of ten or twelve binding posts, a double-pole single-throw switch, a General Electric instrument rectifier, and some five- and ten-megohm resistors arranged as in Fig. 1. These resistors, purchased from a radio store for about ten cents each, were found to vary considerably from their rated values; but this is of small importance as long as their resistances remain constant. A preliminary test indicated that they were very satisfactory in this respect except for a temperature coefficient. It should be remarked here that the whole ensemble may be expected to have a temperature coefficient depending on the particular galvanometer, rectifier, and resistors used. In this case it was of an order of 1 percent per degree.

The full-wave copper oxide rectifier was purchased from the General Electric Company for two dollars. It was labeled as "first made for miniature instruments." Its resistances were approximately as follows: dc side, 200,000 ohms and 3000 ohms; ac side, 100,000 ohms. It was mounted under a glass cover sealed into paraffin to protect it from dust and changes in humidity.

When switch $S$ is closed, the arrangement acts as a dc instrument; when open, as an ac instrument. (It might be better to have a *dpdt* switch here.) Range may be varied by the way the circuit is connected to the binding posts such as $a$ and $b$, $a$ and $d$, $c$ and $h$, $e$ and $f$, etc., also by the addition of a shunt to the galvanometer. As a voltage-measuring instrument it may be calibrated by being connected in parallel with a voltmeter and plotting galvanometer readings *vs* voltmeter readings. Over a part of the range the calibrating curve will be found nearly linear, so that two points are sufficient over this part of the curve.

If one uses a galvanometer arm bearing a telescope and curved scale, a large-scale range is available (25–0–25) which can be read quite accurately. With a total resistor value of approximately 110 megohms, the voltmeter can be used up to about 1000 volts and down to a few millivolts. This allows a resistance of the order of 100,000 ohms per volt.

This device makes possible experiments in alternating currents in which it is desired to measure, for example, the fall of potential across a capacitance of a few microfarads without appreciably affecting the potential difference subtended. The use of an ordinary voltmeter, even of high grade, makes such a measurement unconvincing and unsatisfactory.

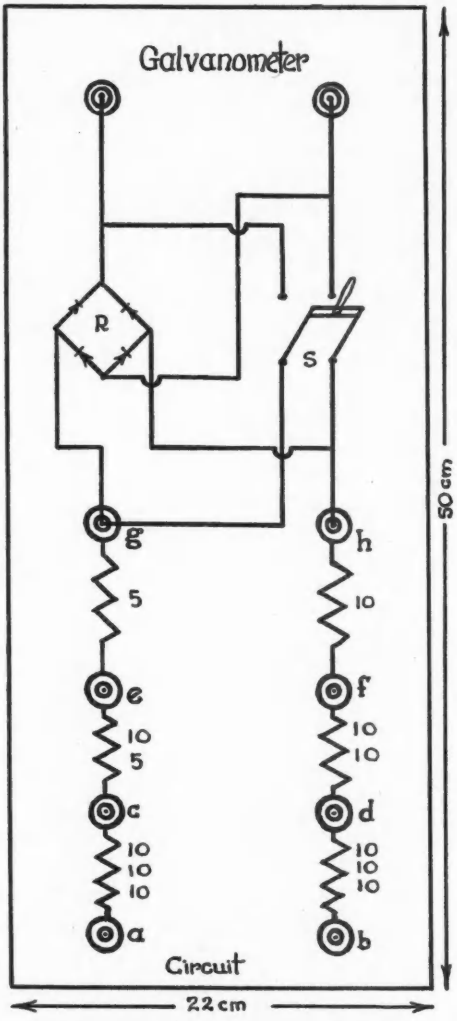


FIG. 1. Wiring diagram of high resistance voltmeter. Here $R$ is a General Electric instrument rectifier. The binding posts $a$, $b$, $c$, $d$, $e$, $f$, $g$, $h$ are connected by 5- and 10-megohm resistors, or combinations of them in series, as shown.

The design, calibration, and use of this high resistance voltmeter have some good pedagogical values. Students have worked previously with the galvanometer and found it excellent for absolute measurements of capacitance and

for use with wheatstone bridges and potentiometers. They are told that ammeters and voltmeters are galvanometers modified for special uses. Here arises a case in point, not just that it *can* be done but that it *must* be done to solve the problem at hand.

This device meets one of the serious criticisms of college teaching, namely, a failure, to some extent, to make students realize that something learned at one point may be of great use at some future unexpected point to solve a totally different problem which might otherwise go unsolved. It also illustrates what we have to say so often about the value of (so-called) "pure science."

## The Use of the Coulomb in Electrostatic Problems

LESTER L. SKOLIL
*University of California, Davis, California*

A SURVEY of a number of well-known college physics textbooks indicates an increasing use of the mks unit of charge (coulomb) in electrostatic problems. Out of more than 30 current physics books examined, about one-third had electrostatic problems in some or all of which the charges involved were expressed in coulombs.

It is desirable for the student in physics to obtain some idea of the order of magnitude of the various physical properties of matter. From solving problems involving these properties, students may gain some idea of relative magnitudes and quite often retain this knowledge for practical application. For this reason, it is desirable that the problems in texts be consistent with practical values.

An examination of the electrostatic problems in physics textbooks shows that some of them give impossible values for the charge on small bodies in air. With the implied configuration, the electric field intensity comes out to exceed the dielectric strength of dry air, a useful value for students to know. The textbooks expressing the value of the charge in coulombs are the ones most often at fault; out of some twenty texts using esu of charge, only one text gave data for problems such that the electric field intensity exceeds the dielectric strength of the medium (approximately 100 dynes/statcoulomb for dry air).

The remainder of the texts examined contained at least some electrostatic problems in which the charge was expressed in coulombs. In all but two of these books, the data in some of the problems gave electric field intensities greater than the dielectric strength of the medium (approximately 30,000 v/cm for dry air).

Some typical problems are given below, without quoting directly from any one text.

1. Calculate the force on a charge in air at a distance of 1 meter from a charge of 1 coulomb.

   The electric field intensity $E$ at a distance of 100 cm from 1 coulomb of charge is $E=9\times10^{11}q/r^2$ $=9\times10^7$ v/cm. Thus, even though the charge was uniformly distributed over a sphere of radius 100 cm, the electric field strength would exceed the dielectric strength of dry air by a factor of 3000.

2. A pith ball of radius $\frac{1}{2}$ cm and a charge of 20 microcoulombs is supported by. . . .

   The electric field intensity at the surface of the pith ball would be over $7\times10^7$ v/cm.

3. A small sphere with a charge of $5\times10^{-5}$ coulomb is located between plates 4 cm apart. . . .

   Assuming a sphere of just under 4 cm, then the electric field intensity at the surface is over $10^7$ v/cm.

4. Three charges of 10 microcoulombs each are located at the corners of an equilateral triangle 2.00 cm on a side. . . .

   Assuming the charges are distributed on maximum size spheres to fit on the triangle, the electric field intensity would exceed $9\times10^6$ v/cm.

A further study of old sets of problems and old examinations used by this writer and by other physics teachers showed similar faults when the charge was expressed in coulombs. It seems it would be well for the physics teacher to examine carefully the electrostatic problems he assigns to make sure the values used and obtained are reasonable and practical.

## An Apparatus for Measuring the Acceleration Due to Gravity

JULIUS H. TAYLOR
*Morgan State College, Baltimore, Maryland*

THE standard method of measuring the acceleration due to gravity by means of the free fall apparatus gives satisfactory results providing one uses an accurate synchronous spark attachment. It is the purpose of this paper to describe a simple elementary laboratory apparatus to measure the constant that can be built with a cash outlay of only a few dollars. The measurements taken with this apparatus are accurate to better than 1.5 percent, thereby providing a satisfactory as well as an inexpensive method of measuring the constant.

The apparatus consists of a disk of dural 0.32 cm thick and 30 cm in diameter Mounted on the periphery of the disk are two brass slugs free to slide along its diameter over a distance of one cm. (See Fig. 1.) The disk can be easily attached to an apparatus similar to the Cenco Rotator[1] by mounting a $\frac{1}{4}$-in. bolt through its center.

As the disk is rotated in a vertical plane, the brass slugs slide back and forth along its diameter until its velocity becomes great enough to cause them to remain in their outermost position. Starting from rest, the velocity of the disk is slowly changed until the slugs no longer possess any motion along its diameter. This velocity is easily disernible because, before it is reached, the slugs make a very loud noise as they slap back and forth in their cradles. The diameter of the disk was so chosen as to give about 75 rpm for the critical velocity. At this

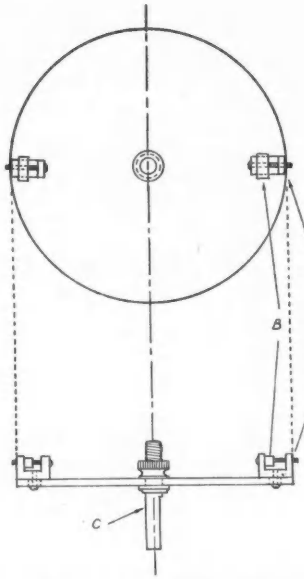


FIG. 1. Disk for measuring the acceleration due to gravity: (A) the cradles, (B) the slugs, (C) bolt for attachment to rotor.

critical velocity, the centrifugal force of the slug in its uppermost position is equal to its weight; that is,

$$m\omega^2 r = mg,$$

or

$$g = r\omega^2,$$

where $m$ is the mass of the slug, $\omega$ the angular velocity, $r$ the radius of the disk, and $g$ the acceleration caused by gravity.

Many data were collected with the apparatus by students on the general college physics level. The results indicated the accuracy to be better than 1.5 percent.

The advantages of this apparatus lie in its simplicity and ease of operation. Its disadvantage is that it involves an understanding of rotary motion which is usually introduced after the student has acquired a through understanding of the acceleration caused by gravity.

[1] Central Scientific Company, Chicago, Catalog No. 74350.

## A Laboratory Examination for General College Physics

W. H. KINSEY AND R. A. RHODES, II
*University of Connecticut, Storrs, Connecticut*

FOR the past 15 years a final examination has been given in the elementary physics laboratories at the University of Connecticut. While this procedure is not at all original with us, it is true, however, that there are some physics departments that do not engage in this practice. We in the department who are involved in the elementary courses are unanimously convinced of the efficacy of the examination and count it 15 percent of the student's grade, while the laboratory reports count only 10 percent.

As a background we shall briefly describe our laboratory as it is run during the semester. Each experiment, which is, incidentally, correlated with the lecture material, is performed and written up by each student during a three-hour laboratory period which comes once a week. The laboratory is worth 1 credit in our 4-credit general physics course. In each section there are 10 setups of apparatus and not more than 20 students. The students, for the most part, work in pairs; and it is the responsibility of the individual instructor to decide whether they have permanent partners of their own choosing or follow a definite rotation scheme. We have reason to believe that a rotation scheme is preferable.

Our reasons for giving the laboratory examination are as follows: (*a*) the student, with the final examination in sight, will take the work more seriously, looking upon it as a part of his course to be understood and remembered rather than as a series of cookbook exercises to be forgotten immediately; (*b*) each student will learn how the apparatus operates and not tend to let his partner do all the work; and (*c*) the results of the final examination will help to divide the students into three levels: (1) those who are good and also have the ability to think, (2) those who have merely learned some of the material, and (3) those who are poor.

The examination itself occupies nine 20-minute periods during the last 3-hour laboratory session of the semester and includes three parts: (1) experimental work consisting of easily worked sections of the experiments performed during the semester, (2) problems on experiments which were done during the semester in which the data are given, and (3) perhaps two pages of true-false and multiple-choice questions on laboratory technique and experiments. Each student works alone, and moves every 20 minutes to a new position according to a schedule on his examination paper.

The experimental questions and problems given are of the usual type found in laboratory finals, but we like to change some of them so that they can be solved easily by the student who understands the principles behind them and knows how to think but will be too difficult for the one who merely memorizes procedures without much understanding or thinking. Examples of our experimental work upon the examination are as follows:

1. *Measurement of resistance of a tungsten filament electric light bulb on a wheatstone bridge.* During the semester, the resistance of a light bulb was measured by a voltmeter-ammeter method under a relatively high temperature condition. In a separate experiment a wheatstone bridge was used, but to measure an ordinary linear resistance. In the examination the student is apt to worry over the inordinately low resistance values of the light bulb obtained by using the wheatstone bridge and might be tempted to falsify his data and results unless he realizes the change of resistance with temperature of the light bulb and has confidence in his measurements.

2. *Efficiency of a pulley system.* The student is asked to plot an efficiency curve with efficiency as ordinate

and load as abscissa. One of the examination questions is to explain the curve. As the load increases, the friction increases so the efficiency curve should decrease with load; but if the student forgets to add the weight of the lower pulley block to the load, the efficiency curve will show an increase with an increase of load. We give the student pulley blocks and weights which will bring out this discrepancy very well.

3. *Mass of a meter stick by the method of moments of forces.* During the semester the student measures in the laboratory the mass of a meter stick balanced at the 50-cm mark with respect to a known mass. In the examination we give him a meter stick which has one end bored with holes, thus making it nonuniform.

4. *Focal length of a positive lens.* The object distance is fixed and is within the focal length. The student cannot, of course, find a real image. He can then employ the pins furnished and use the method of parallax. The thinking student will hold the lens up so that another distant object will give an image approximately at the focal point. Then he can check his computed value of focal length.

The results of the examination over the past four semesters have been analyzed and give the following information. First, there is far greater correlation between a student's grade in the final laboratory examination and the grade earned in the rest of the course, exclusive of the laboratory reports and laboratory examination, than between the laboratory examination and laboratory report marks. The chief reason for this is that a poor student can be carried along by a good partner and turn in ostensibly good work, particularly if he has the same partner throughout the semester. Secondly, we investigated the grades of the past two years and found that the students who answered the "thinking" questions correctly were in general the "*A*" and "*B*" students of the University, while those who missed the "thinking" questions were the "*C*" and "*D*" students.

Thus, we conclude from our experiences that the laboratory examination is a better measure of the ability of a student than the laboratory reports. In lieu of some yet undeveloped way of improving the laboratory report requirements, laboratory examinations should be given and counted toward the course grade.

## A Simple Demonstration of Coriolis Force*

Arthur A. Klebba and Henry Stommel
*Woods Hole Oceanographic Institution, Woods Hole, Massachusetts*

SEVERAL years ago, in trying to build a simple teaching aid for demonstrating Coriolis forces, we devised the following apparatus. A twelve-inch diameter pan was fastened to a phonograph turntable, and filled, while rotating, with molten paraffin, which formed a paraboloid of revolution upon cooling. The surface of the paraboloid was rather rough at first but was easily made true by repeated melting and solidification of the top surface by means of a heat lamp. A Dove prism was mounted inside a ball-bearing race and rotated by a small variable speed motor at exactly one-half the angular velocity of the turntable. Under these conditions, when the prism assembly was placed above the turntable so that the two axes of revolution coincided, the turntable as viewed through the prism (Fig. 1) appeared to stand still.

The paraboloid is a surface normal to the resultant of the centrifugal and gravity forces, so that a small marble

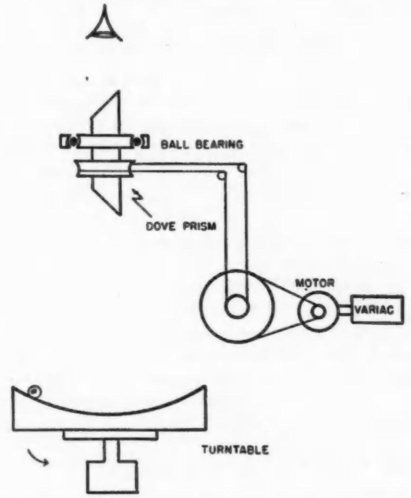


FIG. 1. Apparatus for viewing motion of objects on rotating surface.

remains at rest at any point on the surface. If displaced by an impulse the marble executes inertial motions[1] which can be clearly seen through the prism. For example, it is easily seen that the marble always moves in a circle relative to the paraboloid with twice the angular velocity of the pan.

The apparatus should be particularly helpful in teaching physical meteorology and oceanography, because it is frequently difficult to impress students with the reality of Coriolis forces. Various modifications of this apparatus to demonstrate other elementary phenomena on a rotating coordinate system will be obvious to the teacher.

* Contribution No. 545 from the Woods Hole Oceanographic Institution.

[1] Brunt, D. *Physical and Dynamical Meteorology* (Cambridge University Press, London, 1941), p. 167.

# Letters to the Editor

## A Demonstration of Bernoulli's Principle

A VERY satisfactory demonstration in connection with Bernoulli's principle consists in directing a jet of compressed air across the top end of a short length of ordinary glass tubing which extends vertically downward into a beaker of water. This simple arrangement produces a very effective spray, and lends itself well to explanation.

E. SCOTT BARR

*University of Alabama,*
*University, Alabama*

## Has Pressure Direction?

IN several preceding papers there has been discussed the question whether the hydrodynamic pressure is a scalar or a vector.[1] In fact neither is true; it is actually a tensor. This is best understood remembering that among other possibilities an ideal fluid can be defined by the conditions that; (1) the tangential tension is identically zero, (2) the tension cannot be positive (i.e., the pressure cannot be negative), and (3) the compressibility is zero. Condition (1) means that in the pressure tensor (i.e., the negative tension tensor) all nondiagonal elements are identically zero, which implies that all diagonal elements are equal. In this case the pressure tensor is fully defined by one magnitude only; it then can be represented by a scalar. So the pressure in hydrodynamics has no directional properties whatever. For more details see any textbook on the subject.[2]

MICHAEL DANOS

*Institut für Theoretische Physik,*
*Heidelberg, Germany*

[1] A. R. Stickley, Am. J. Phys. **18**, 322 (1950); R. D. Summers, *ibid.* **14**, 311 (1946); *ibid.* **17**, 319 (1949); V. P. Barton, *ibid.* **17**, 318 (1949).
[2] For instance: A. Sommerfeld, *Vorlesungen über theoretische Physik*, (Leipzig, 1949), p. 54.

## A Mercury Light Source

SEVERAL economical types of mercury light sources have been suggested for the beginning laboratory. Schwinn[1] has suggested the use of a type 816 mercury-vapor rectifier. Kirkpatrick[2] suggests an ordinary fluorescent lamp while McCay and Bishop[3] use a standard germicidal lamp.

Another satisfactory source may be constructed with parts usually found in a physics laboratory plus three items that may be purchased for less than $15.00. These items are a GE S-4 sunlamp which operates from a (GE No. 58G720) ballast, and the special socket for the lamp. The other parts consist of a tin can for the lamp house, a ring stand to hold the lamp at the desired height, and a small metal cabinet to house the ballast. An off-on switch may be mounted on one side of the cabinet. It may also prove convenient to have an outlet on the same side of the cabinet which will permit the lamp to be disconnected. A filter holder may be constructed of plywood or some other material and clamped to the lighthouse with a coil spring.

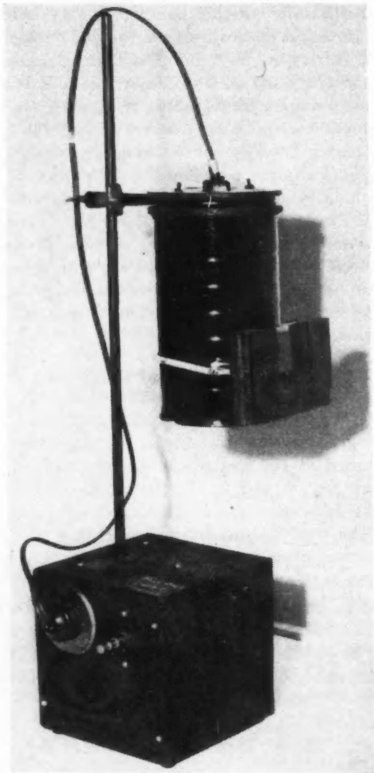

FIG. 1. Mercury light source.

At William Jewell College we have constructed several of these units, one of which is shown in Fig 1.

WALLACE A. HILTON

*William Jewell College,*
*Liberty, Missouri*

[1] M. W. Schwinn, Am. J. Phys. **15**, 279 (1947).
[2] P. Kirkpatrick, Am. J. Phys. **15**, 359 (1947).
[3] M. S. McCay and E. S. Bishop, Am. J. Phys. **16**, 361 (1948).

## Lissajous Figures in Melde's Experiment

THE standard experiment on a vibrating string, known familiarly as Melde's experiment, lends itself nicely to a demonstration of the classical Lissajous figures.

The usual procedure needs little recitation. The vibrator or fork is so supported that the vibration of the prong is

either along the line of the string or normal to it, the purpose being to provide essentially a "plane-polarized" vibration of the string. It is invariably noted, however, that the motion of the string is seldom ideally in one plane, for this requires an *exact* loading and a *precise* orientation of fork and string and pulley. What usually happens is that the string, at the loops, executes either elliptic or circular motion, the departure from one plane being hardly noticeable, however, if the adjustment is only a little short of critical.

It occurred to me that this departure from motion in one plane would be enhanced by further lack of alignment of fork and string, and that vibrations in this fashion, being, so to speak, a combination of the two standard modes, might lead to Lissajous figures. *This is precisely what ensues.*

If the fork is put obliquely to the line of the string the loops in the vibrating string execute no simple pattern, as is the case in the two standard positions. The geometry realized at the loops constitutes a beautiful demonstration of Lissajous figures. The demonstration may be accompanied by an analytic discussion of the generation of Lissajous figures resulting from a simultaneous superposition of simple harmonic motions.

Stroboscopic illumination reveals a further most interesting detail. A migration of the nodes arises in this oblique arrangement, that is, a node "travels" from its extreme position for one orientation of string and fork (parallel, say) to its extreme position for the other orientation (normal), the distance of this migration being exactly that between two successive loops. This might well be expected since the number of loops (for the same tension) obviously changes in the two standard positions. It is to be recalled that with the fork vibrating in the direction of the line of the string the frequency of the string is one-half that of the fork.

Since reporting these things to the extent above, my attention has been called to earlier discussions of the phenomena so that these observations are not altogether original with me, as I first believed. We should probably say with Ecclesiastes: ". . . and there is no new thing under the sun. Is there a new thing whereof it may be said, See, this is new?"

In a note to *Nature*, November 4, 1909, C. V. Raman reported on the motion of the nodes of a vibrating string. In 1911[1] he extended the discussion and examined the motion which these "fictitious nodes" execute parallel to the string, the range of motion being equal to the whole length of a loop. The string is attached to the prong of the fork "so that it lies in a plane perpendicular to the prongs but in a direction inclined to their line of vibration. Under these circumstances the motion of the prong may be resolved into two components, one perpendicular and the other parallel to the string. The two oscillations can be made to occur in perpendicular planes." In a very elegant and poetic fashion Raman describes the observation: "The surface generated by the moving string is one of great delicacy and purity, and an adequate idea of it can only be had on actually performing the experiment."

Later papers by Raman[2,3,4,5] extend the discussion both theoretically and experimentally. Excellent photographs accompany them all. The exposition is beautiful and indeed, exciting to read. "A very neat modification of Melde's experiment can be arranged. . . ." He speaks of the "great beauty of the form of oscillation," and refers to the oscillation mechanism as "one of the most fascinating problems in the dynamics of vibrating bodies."

In 1917 Jones and Phelps published "Notes on Melde's Experiment"[6] in which they consider both the "parallel" position and the "transverse" position of string and fork. The "parallel" position is that with which Raman was principally concerned. "For certain tensions the paths of the elements of the string may become circles, in planes perpendicular to the length of the string." They show an array of beautiful photographs taken stroboscopically from various angles.

I recommend this inquiry as fascinating to pursue. I am indebted to Mr. G. S. Bennett for directing my attention to it. I would not have unearthed this little treasure otherwise.

Julius Sumner Miller

*Dillard University,*
*New Orleans, Louisiana*

1 C. V. Raman, Phys. Rev. **32**, 309 (1911).
2 C. V. Raman, Phys. Rev. **35**, 449 (1912).
3 C. V. Raman, Phys. Rev. **4**, 12 (1914).
4 C. V. Raman, Phys. Rev. **5**, 1 (1915).
5 C. V. Raman, Phil. Mag. **24**, 513 (1912).
6 A. T. Jones and M. E. Phelps, Phys. Rev. **10**, 541 (1917).

AMERICAN JOURNAL OF PHYSICS VOLUME 19, NUMBER 4 APRIL, 1951

# RECENT MEETINGS

## Chicago Section

The Chicago Chapter of the AAPT held its fall meeting Saturday, December 2, 1950, at the University of Chicago. Dr. Harold R. Voorhees was the host. Twenty-seven members and guests were present.

In the morning, the betatron and cyclotron laboratories of the University of Chicago were inspected; Robert F. Dvorak, health physicist, acted as guide. Finishing touches are being put on the new cyclotron, and test runs will be made shortly. Following luncheon in Hutchinson Commons, a business meeting was held in which the following were elected as officers for the next year:

*President*—Clarence J. Overbeck,
*Vice President*—W. W. Colvert,
*Secretary*—Robert M. Besancon.

Illustrated talks were presented by Paul F. Copeland on **Pool-type Cathodes,** Ora L. Railsback on **The Physical Basis of Piano Tuning,** and Russel A. Fisher on **Recent Work in Infrared Spectroscopy at Northwestern University.**

William R. Anderson, *Secretary*

## Kentucky Section

The annual fall meeting of the Kentucky Section of the American Association of Physics Teachers was held on November 11, 1950, in Science Hall, Berea College, Berea, Kentucky. The meeting was attended by thirty-five members and guests.

The program of contributed papers is given below. Following the papers the group had an excellent luncheon in the College Boarding Hall.

**An integrating course for the physics major.** CARL E. ADAMS and RALPH A. LORING, *University of Louisville.*

**Laboratory guide for statics and dynamics.** EARLAND RICHIE, *Centre College.*

**Experiment demonstration to determine rifle bullet velocity.** DONALD WORTH, *Berea College.*—An arrangement for general laboratory determination of the velocity of a .22 rifle bullet was shown and discussed as to accuracy. Basically the setup consisted of a variation of the rotating disk method. Two disks were used, each composed of wrapping paper 24 inches in diameter, clamped between 16-inch corrugated board disks in turn clamped between smaller wooden disks. These disk assemblies were mounted on the arbors of a bench grinder, the grinder motor being elevated on a small table so that the bullet would pass beneath it. The rifle was rigidly mounted in a horizontal position accurately parallel to the motor shaft, eliminating skew shots. By use of a plumb line or direct superposition of the two disks, the relative rotation of the bullet holes may be found to within three percent. The students measure the chordal distance between radial lines through the bullet holes at the points these intersect a reference circle of known radius. From this information the relative angle of rotation is obtained. Since the grinder motor is of the repulsion-induction type, line voltage variations have no effect. The present over-all accuracy of determination is somewhat less than five percent.

**The tracing of skew rays by analytical methods.** CHARLES C. DALTON, *University of Kentucky.*—Considering only homogeneous media, one may say that an optical element is a medium having an index of refraction which may differ from the indexes of the media adjoining it and has the property of altering the form of the wave front in the path of which it is placed. The normals to the advancing wave front are called rays, and a description of the changing form of a wave front is called ray tracing. Ray tracing is therefore employed either to describe the process of image formation or to describe spectroradiometrically the effective radiant energy density at points in space.

A skew ray is a ray which in transit through an optical element is such that its incident, internal, and emergent directions are not coplanar—a condition due to the fact that the normal to the second refracting surface at the second refracting point is not contained in the plane determined by the incident ray and the normal to the first refracting surface at the first refracting point.

The author's solution to the skew ray problem, using the principles of solid analytic geometry, has the advantage of being applicable to elements, the refracting surfaces of which consist of any configurations whatever.

**Sonic determination of air temperature.** F. D. WATSON and K. O. LANGE, *University of Kentucky.*—The need for sonic thermometry arises in meteorological investigations at high altitudes. In the rarefied air, possessing low thermal conductivity and reduced heat capacity, radiation effects become more pronounced. Consequently, conventional thermometric elements will obtain thermal equilibrium at temperatures other than that of the air.

Inasmuch as the velocity of sound in air is essentially a function of temperature and composition of the air, it has been possible to determine air temperature sonically. A pulse system for the measurement of the velocity of sound was constructed. This system employs a triggered electronic circuit for measuring the time of transition of a sound pulse over a fixed path of approximately two meters. In its present form the circuit produces a time-indicating sweep pattern on a cathode-ray tube screen.

Laboratory tests indicate that the system should be applicable up to altitudes of approximately ninety thousand feet. With a correction for the effects of water vapor, the over-all accuracy is $\pm 1°C$. By telemetering information to a ground station it is possible to reduce the weight of the airborne unit to about 3.5 lb, which is practicable for balloon soundings. It is thus possible to obtain temperature measurements free from errors due to radiation effects or thermal lag.

**An investigation of gas amplification in a proportional counter.** J. E. HOPSON, *University of Kentucky.*—M. E. Rose and S. A. Korff (Phys. Rev. **51**, 850 (1941)) have obtained theoretically an expression for gas amplification in a proportional counter of cylindrical geometry.

A proportional counter was constructed in which it was possible to vary the gas, the pressure, wire radius, cylinder radius, and operating voltage. Special arrangements were made to minimize end effects. Using a polonium source of ionization, a no-gain preamplifier, a decade attenuator, a linear amplifier, and a scalar, twelve curves of gas amplification *vs* operating potential were obtained for methane, each for a different set of counter dimensions or gas pressure.

Theoretical curves were then fitted to the experimental data. In each case the predicted threshold was too high, but the theoretical curves had the correct shape between amplification factors of 15 and 3000. However, the dependence of amplification on wire size was the only theoretical expression checking with the data within the limit of experimental error.

Curves were also obtained for operating potential *vs* pressure at constant amplification. The Geiger region was determined; and the pulse-height spread was determined at the proportional threshold, in the proportional region, and in the Geiger region for one experimental curve.

**Radioactive ore in Boyle and some adjoining counties.** ROY ELLIS, *Centre College.*

**The University of Kentucky van de Graaff generator.** T. M. HAHN, JR., *University of Kentucky.*—The University of Kentucky van de Graaff generator has been altered so that variable electrostatic focusing could be installed. A new ion source has been installed, and through a system of selsyns it is possible to vary the operating voltages of the

ion source while the generator is in operation. A 90° deflecting-analyzing magnet is being installed to convert the generator to horizontal operation, and also, through a corona feedback circuit, to provide stabilization of the operating voltage of the generator. With the modified ion source, total beam currents of as high as 100 microamperes have been obtained.

This work has been accomplished through the joint efforts of the Physics Department of the University of Kentucky and the NEPA Project of Fairchild Aircraft Corporation.

**A modified Wilson cloud chamber.** GEORGE C. PATTERSON, *University of Kentucky.*

RICHARD HANAU, *Secretary*

## Physics Club of Philadelphia

During the year 1949–1950, the Physics Club of Philadelphia entertained four lecturers with an average attendance of seventy-five members and guests.

DR. PHILIP MORRISON, *Cornell University.* **The properties of mesons and nucleons today.** (October 21, 1949).

DR. ALBERT C. WALKER, *Bell Telephone Laboratories.* **Growing piezoelectric crystals.** (December 7, 1949).

DR. ROBERT C. GORE, *American Cyanamid Company.* **Infrared spectrometry and the reflecting microscope.** (March 17, 1950).

DR. CHARLES P. BAZZONI, *Sun Oil Company.* **Geophysics in the oil industry.** (May 5, 1950).

The meeting of December 7, 1949, was held as a joint meeting with the Franklin Institute; the meeting of May 5, 1950, was a dinner meeting which concluded with the election of officers for the year 1950–1951.

In addition to the four scheduled meetings listed, the Physics Club joined with the Franklin Institute and other science groups in the Philadelphia area in the formation and operation of the PHILADELPHIA SCIENCE COUNCIL. The Science Council was organized to make a search among each Philadelphia and Suburban High School graduating class for science-talented youth. Selections have been made on the bases of recommendation of high school teachers, a comprehensive examination, and personal interview. At the Science Council Award Dinner which was held June 4, 1950, student memberships to the Franklin Institute were presented to the ten high school seniors showing the greatest interest and attainment in physics. It is expected that the number will be increased for the current year.

Participation in the activities of the Science Council has shown the need of special physics laboratories for the use of high school seniors showing special promise in physics. At the present time, studies are being made in two directions, first the possibility of an adequate central laboratory operated by the school district, and second, the possibility of assignment of space in college physics laboratories for the use of high school seniors after school hours and on Saturdays. Some plans have been initiated by one of the Philadelphia universities, but on only a limited scale.

The officers elected for the year 1950–1951 are: President, G. WILLIAM DONOVAN, *Murrell Dobbins Vocational School;* Secretary, MABEL A. PURDY, *Kensington High School;* Executive Committee, FRANK P. HOCHGESANG, *Socony Vacuum Laboratories,* and CARL T. LEANDER, *School of Dentistry, University of Pennsylvania.*

MABEL A. PURDY *Secretary*

## Oregon Section

The 55th meeting of the Oregon Section was held on November 11, 1950, at Washington State College, Pullman, Washington, DR. M. A. STARR, *University of Portland,* presiding. Guided tours of the new Technology Building were conducted by members of the local college staff, and demonstrations and explanations were given of the numerous research projects currently being conducted. The formal program included the following papers.

**Field emission at a billion amperes per cm².** J. K. TROLAN and W. P. DYKE, *Linfield College.*

**An improved electron projection microscope.** FRANK GRUNDHAUSER, *Linfield College.*

**The detection of the neutral meson.** LAWRENCE S. GERMAIN, *Reed College.*

**The origin of the lighter elements.** WILLIAM BAND, *State College of Washington.*—Maria Mayer and Teller have proposed accounting for the relative abundances of the heavier elements as products of the fission of a cold polyneutron fluid. The present paper proposes to account for the lighter elements as having been formed at an earlier stage in evolution when temperatures were high enough to permit nuclear reactions, and the primordial neutron-rich nuclear fluid first separates into two phases: the condensed super-clusters or stars which eventually cool to become the Mayer-Teller polyneutrons, and an atmosphere or vapor of small clusters that become the lighter elements.

The quantum statistics of a dissociative clustering nucleon gas is worked out and yields equilibrium abundances of the small clusters or light nuclei in the vapor that agree qualitatively with the observed abundances if the temperature $T$ of the equilibrium is assumed to be such that $kT=5$ Mev.

**Note on an inexpensive strong U-V source.** FRED W. DECKER, *Oregon State College.*—A type-RS sunlamp can be dismantled to provide a source of the mercury U-V spectrum. The glass envelope is removed to expose the quartz mercury discharge tube. The thermal switch and ballast lamp filament are replaced by a manual switch and a suitable wire resistor. Power dissipation of the quartz discharge tube is about 80 watts. Wavelengths below 2500A are included in the output. The total cost is approximately ten dollars for the lamp and all parts.

FRED W. DECKER, *Secretary*

# ANNOUNCEMENTS AND NEWS

## Book Reviews

**Introductory Nuclear Physics.** DAVID HALLIDAY. Pp. 558, Figs. 282. John Wiley and Sons Inc., New York, 1950. Price $6.50.

The hazards one must face in writing a textbook on the fluid and expanding subject matter of nuclear physics have been so great that, notwithstanding the importance of the field and the intense activity within it, the number of such books has been disproportionately small. Those which have appeared often omit segments of subject matter quite properly within the domain of nuclear physics and, in a few instances, they can only be characterized as spotty in their coverage. When viewed against this background, the accomplishment of Professor Halliday must be regarded as indeed heroic.

The book is primarily intended for a beginning graduate course; with the omission of sections in finer print, it can serve as a text for an advanced undergraduate course. Professor Halliday assumes that the student has a qualitative understanding of wave functions and quantum numbers, such as that obtained in a conventional undergraduate treatment of modern physics. Nevertheless, first references to quantum ideas are accompanied by brief introductory remarks which assure that a bright sophomore or junior physics major who happily chances to peruse the book will not likely be overwhelmed by too abrupt injection of new material. At the other end of the spectrum of physicists, the accomplished specialist in some phase of nuclear physics will find quite to his liking the immense number of topics covered and the well-chosen literature references which fulfill Professor Halliday's stated hope that the text will also serve as a first source book of methods.

A major problem in the beginning graduate treatment of nuclear physics lies in the free use of results from quantum mechanics and relativity before the student has met systematic treatment of these subjects. At worst, such cart-before-the-horse procedures, which are of necessity common in undergraduate and early graduate courses in modern physics, torment the student with vast and intricate mazes of seemingly unrelated facts and ideas. At best, these procedures not only present organized subject matter, but they also fire the student with enthusiasm for studies which lie ahead. Professor Halliday's clear style and his abandonment of the historical development in favor of a logically organized approach surely combine to avoid confusion. His modern point of view and his copious references to recent literature, both in the text and in the many excellent problems, enable the student to reveal to himself an inspiring panorama of present-day nuclear physics. To the reviewer, this skillful and effective use of recent literature (40 percent of the illustrations are from publications after 1945) is one of the two outstanding features of the book.

The second is completeness. Although some nuclear specialists might wish greater emphasis on their favorite topic, the reviewer is confident that most physicists will find Professor Halliday's a remarkably balanced treatment of most of the topics of nuclear physics. The student of this text will, on one hand, learn something about the description of nuclear scattering through the phase-shift parameters and, on the other hand, he will have accurate impressions, based upon prominently displayed tabulations, of the physical size, the electric power consumption, and other data concerning the apparatus with which scattering experiments are carried out. He will learn more about beta-ray spectrometers than merely that charged particles are focused in magnetic fields; for example, he will note that the magnetic field of a beta-spectrometer is one or two orders of magnitude smaller than that of a cyclotron. Such "engineering" information is not to be scoffed at by even the longest-haired theoretician, for a useful and reliable physical intuition rests solidly upon an accurate feeling for orders of magnitude. And surely the experimentalist, however grimy his hands, can have only good words for Professor Halliday's discussions of parity, of nuclear shell structure, and of the Breit-Wigner theory of nuclear resonances. The lifetime of $\pi$-mesons, the mass distribution of fission fragments, the partial pressures of the gases in a typical Geiger tube, the rise- and delay-times of a linear amplifier—this book has something to say on nearly every topic which should be of current interest to the beginning graduate student.

Finally, a particular feature of the book so pleases this reviewer that he can no longer contain his delight. He has often been at a loss to answer the eager new graduate student who, having seen any recent physics periodical, asks to be referred to a single comprehensible book indicating what electric quadrupole moments are and how nuclear quadrupole moments are measured. Professor Halliday has provided an excellent answer for this question as well as for most others which such students can ask about nuclear physics.

G. E. PAKE
*Washington University*

**Albert Einstein—Philosopher-Scientist.** Edited by PAUL ARTHUR SCHILPP. With contributions by Einstein, Sommerfeld, de Broglie, Rosenthal-Schneider, Pauli, Born, Heitler, Bohr, Margenau, Frank, Reichenbach, Robertson, Bridgman, Lenzen, Northrop, Milne, Lemaitre, Menger, Infeld, von Laue, Dingle, Gödel, Bachelard, Wenzel, Ushenko, and Hinshaw. Pp. 781. The Library of Living Philosophers, Inc., Evanston, Illinois, 1949. Price $8.50.

A great mind scarcely ever has been a specialist. Almost all of modern physics has been shaped by Albert Einstein

in such a way that it would be difficult to imagine what would have been the development of either quantum or relativity theory without him.

Many branches of physics were developed by him: statistical mechanics, theory of specific heats, Brownian motion, quantum theory of the photoelectric effect, emission and absorption of radiation, Einstein-Bose statistics, Einstein-de Haas effect, special and general theory of relativity, cosmology, unified field theories, philosophy of science. We can touch only relativity in this short review.

The special theory of relativity has been so completely integrated into every day's work in physics that we do well to be reminded of the problems of early days. To quote Einstein about his colleagues' contributions, "Before Minkowski's investigation it was necessary to carry out a Lorentz transformation on a law in order to test its invariance under such transformations; he on the other hand succeeded in introducing a formalism such that the mathematical formulation of the law itself guarantees its invariance under Lorentz transformations." The main achievements of special relativity theory are the discovery of the relativity of simultaneity, the energy-momentum conservation law (mass-energy relation), and de Broglie's wave mechanics, a direct consequence of relativity theory applied to long-established quantum principles.

General relativity theory is still more exclusively Einstein's own work. He was the first to recognize that our actual space is non-euclidean. Thereby, he achieved a derivation of the newtonian theory of gravitation from nothing but the hypotheses of the principle of general covariance and for the assumed (but experimentally well-established) equivalence of inert and heavy mass; all this without the *ad hoc* assumptions which Newton had to introduce, such as that of the mutual attraction of masses and that of the existence of absolute space. The latter assumption contains unsurmountable contradictions in itself. Besides, the three well-known deviations of general relativistic physics from classical physics were predicted by Einstein.

A century ago C. F. Gauss, Bernhard Riemann, and Felix Klein realized that the geometry of a space is characterized by the group of transformations which leaves the geometry invariant. This discovery underlies all of modern physics. In developing the general theory of relativity, Einstein recognized that there is an almost compelling necessity with which the transformation group determines the field laws. In general relativity, this is the group of general coordinate transformations under which the field laws are supposed to be covariant.

The group of general relativity demands that the simplest invariant laws be no longer linear or homogeneous in the field variables. This fact led to another profound discovery: In the relativity theory of gravitation, the law of motion (geodetic line) was originally postulated independently in addition to the field equations—just as it always was in electromagnetic or other field theories. In 1928 it became apparent that the law of motion need not and must not be assumed independently, but that it is already implicit contained within the law of the gravitational field.

This book has a refreshing spirit, full of pep and controversy, with a lot of humor slyly hidden away in some side remarks and cracks. A complete bibiliography of Einstein's works is given by Margaret Shields.

Albert Einstein is, perhaps more than any of us, a human being, little understood, full of spirit and courage, a real friend. He believes in freedom and likes the atmosphere of unprogrammed, independent research. "It is, in fact, nothing short of a miracle that the modern methods of instruction have not yet entirely strangled the holy curiosity of inquiry, for this delicate little plant, aside from stimulation, stands mainly in need of freedom, without this it goes to wreck and ruin without fail. It is a very grave mistake to think that the enjoyment of seeing and searching can be promoted by means of coercion and sense of duty." But surely his concern for individual freedom is not the only core of his message.

In his autobiographical notes, Einstein writes that he felt himself called upon to think and search after truth in the realm of nature rather than to act and to enter into problems of human relations. This he says because he is humble. Certainly he suffered and he did what he was able to do to relieve human suffering, unafraid of being despised and persecuted for telling the unpopular truth. Recently he wrote to the Society for Social Responsibility in Science: "The problem of how man should act if his government prescribes actions or society expects an attitude which his own conscience considers wrong, is indeed an old one. It is easy to say that the individual cannot be held responsible for acts carried out under irresistible compulsion, because the individual is fully dependent upon the society in which he is living and, therefore, must accept its rules. But the very formulation of this idea makes it obvious to what extent such a concept contradicts our sense of justice. External compulsion can, to a certain extent, reduce, but never cancel the responsibility of the individual. In the Nuernberg trials this idea was considered to be self-evident. Whatever is morally important in our institutions, laws, mores, can be traced back to interpretation of the sense of justice of countless individuals. Institutions in a moral sense are impotent unless they are supported by the sense of responsibility of living individuals."

We think that we cannot conclude this personal note better than by repeating his own words of some 25 years ago: "Very different are the people and the motives of the souls of those who do scientific work. Some carry on their studies with the contented feeling of the brillancy of their minds; for them science is a game offering powerful experience of satisfied ambition. There are also many who but for profitable aims risk some headache. If an angel were sent by God to drive out of the temple of science all those people, there would be embarrassingly few left. Among them would be our Max Planck."

Herbert Jehle
*University of Nebraska*

**Fundamentals of Acoustics.** LAWRENCE E. KINSLER AND AUSTIN R. FREY. Pp. 516+vii. Figs. 170. $5\frac{5}{8}\times 8\frac{5}{8}$ in. John Wiley and Sons, Inc., New York, 1950. Price $6.00.

This reviewer has felt for some years that there existed no truly suitable text for an intermediate acoustics course. Available American books were either ridiculously elementary or altogether too sophisticated, while a couple which were of about the right level (British in origin) were rather sadly outdated. Such topics as the phonodeik, the bird call, etc., scarcely deserve any particular attention in a present day acoustics course. The impetus given to the science of acoustics by the 1941–1945 war made the lack more keenly felt. It was with considerable hope and enthusiasm, then, that the present book—the first new acoustics book in a long time—was received.

The preface states that the book is aimed at advanced undergraduate and graduate students, and further, that it was developed from a course given to students of engineering electronics. The authors state that the first nine chapters may be used as a one-semester course in theoretical acoustics, while the remaining seven chapters take up various applications. The level of the mathematics involved is such that students in the above category should be able to handle it. At the same time, mathematical treatments are, in general, sufficiently complete that a good student having had general physics and calculus should, by dint of hard work, be able to follow the developments. The exponential formulation for periodic quantities is used throughout, and some familiarity with differential equations and circuit theory—particularly the impedance concept—would be of great value to the student.

The first nine chapter headings are: Simple Harmonic Motion, Vibrating Strings, Vibrations of Bars, Circular Membranes and Plates, Acoustic Plane Waves, Transmission Phenomena, Spherical Acoustic Waves, Resonators and Filters, and Absorption of Sound Waves. The treatments are, in general, the usual ones. In the first chapter, techniques for solving the SHM and damped oscillator equations are examined in some detail, and for the partial differential equations encountered in chapters 2 to 4 the solutions are obtained by separation of variables. The effects of boundary conditions of different kinds are worked out explicitly and completely. A minor complaint enters here, in the treatment of strings, in that the fourier series technique is assumed to be known to the student. The technique is introduced on p. 51 with the words "applying the fourier theorem to equation 2.22 gives

$$A_n=(2/l)\int_0^l y_0(x)\sin(\omega_n x/c)dx,\text{"}$$

with no statement of any sort as to what the fourier theorem is. It would seem that this would be an excellent place to devote a page or so to the fourier expansion. Hyperbolic functions and bessel functions are introduced when needed, the nature of these functions examined to some extent, and values tabulated in the appendix.

The chapter on transmission phenomena treats passage from one medium to another, through three media, including the nonreflecting or impedance matching layer, and transmission across changes of area in pipes. The chapter on spherical waves, among other things, treats the piston source in detail—effect of an infinite baffle, directivity pattern and directivity criteria, radiation impedance, etc. Resonators are treated in chapter 7 in a conventional manner, and the standard electromagnetic electrical analog introduced. No mention is made of the electrostatic analog. The chapter on absorption treats the damped wave equation, viscous absorption, and the effects of heat conduction, porous solids, and briefly mentions molecular absorption. This reviewer is very disappointed in the latter treatment, both in extent and intensity, since he feels the development and verification of this concept represents one of the great advances in acoustics.

The last seven chapter headings are: Direct Radiator Loudspeakers, Horn-Type Loudspeakers, Microphones, Psychoacoustics, Architectural Acoustics, Underwater Acoustics, and Ultrasonics. The chapters on loudspeakers and microphones analyze various types and treats design details to somewhat greater extent than would be expected in a course for physics majors, thereby reflecting the engineering slant which is manifest all through the book. Reciprocity calibration manipulation is described in some detail, although the discussion of the theoretical background of the technique leaves something to be desired. The chapter on psychoacoustics, which this reviewer is least qualified to comment upon, treats largely of the mechanism of the ear and theories of hearing. The chapter on architectural acoustics discusses Sabine's equation (of course!) and also shows how the wave analysis leads to Sabine's equation in the limit, as well as introducing the acoustic impedance concept for wall materials. The last chapter, on ultrasonic phenomena, mentions briefly some of the more spectacular effects, the interferometer, the Debye-Sears light-diffraction technique, and a rather more than usually detailed discussion of piezoelectricity.

The chapter on underwater acoustics has been taken out of order and given a paragraph of its own, because the reviewer feels that this chapter is the outstanding feature of the book. Obviously, here the authors are completely at home—indeed, the statement that the authors are connected with the U. S. Navy is superfluous after one has read this chapter. Here are descriptions of the transmission anomaly, Sonar, Sofar (the deep sound channel), and other aspects of acoustics of particular interest underwater, and of great growth during the late war. To the reviewer's knowledge, this chapter represents the first up-to-date, easily available discussion of these important phenomena.

A goodly number of problems is given for each chapter, the numbers ranging from 9 to 18. While some of the problems appear to be tolerably difficult derivations, others are numerical computations (although not necessarily easy), so it would seem that problems could be selected to suit the class. No particular search was made for typographical errors, but the typography seems clean, with the exception of Fig. 3.3 (b) for $n=3$, in which there has been an obvious vertical off-set of nodal pattern and bar.

In summary this is an excellent book, both as a reference and as a text. Any person concerned with acoustics should have a copy. As a textbook it would be of particular value

at the level indicated for a class of engineering bent. This reviewer does not feel that the hiatus in acoustics texts mentioned in the first paragraph has been completely removed, particularly for a junior grade course, but he is of the opinion that this book is the best text on acoustics currently on the market.

G. S. BENNETT
*Michigan State College*

## University of Alabama Physics Building Dedication Meeting

On November 6, 1950, the University of Alabama had a dedication program for its new Physics Building which was occupied by the physics department at the beginning of the fall semester. The new building has about 50,000 square feet of floor space that is divided into lecture rooms, class rooms, laboratories, and individual research rooms. One wing forms the astronomical observatory, which is equipped with a new 10-in. refractor.

The dedication program consisted of lectures on various timely topics in physics and astronomy. The list of speakers and their subjects is as follows: for the morning session, ROSE C. L. MOONEY, *Newcomb College of Tulane University*, **Modern Developments in X-Ray Structural Study,** and ARTHUR E. RUARK, *The Johns Hopkins University Institute for Cooperative Research*, **Physics and Freedom;** luncheon speaker, ALVIN M. WEINBERG, *Oak Ridge National Laboratory*, **Atomic Energy and Education in the South;** afternoon session, WILLIAM V. HOUSTON, *The Rice Institute*, **Recent Developments in Superconductivity,** and KARL K. DARROW, *Bell Telephone Laboratories*, **Electricity in Metals and Semiconductors;** and for the evening session, DONALD H. MENZEL, *Harvard College Observatory*, **Why Take the Sun for Granted?**

The ladies of the physics department served tea during open house from 4:30 to 6:00 P.M. At the open house period following the evening lecture, the observatory was open to permit visitors to look through the telescope.

An added feature which attracted great attention was the exhibit from the Oak Ridge Institute of Nuclear Studies. Mr. David L. DeJarnette, who is director of the Institute's museum on atomic energy, brought the exhibit.

Approximately two hundred people from out of town came to the program.

ERIC RODGERS

## New Members of The Association

The following persons have been made members or junior members (*J*) of the American Association of Physics Teachers since the publication of the preceding list [Am. J. Phys. 19, 121 (1951)].

Agresta, Joseph (*J*), 32–22 34 St., Long Island City 6, N. Y.
Almond, Harold Briggs (*J*), Physics Dept., Sterling Hall, University of Wisconsin, Madison, Wisc.
Amar, Henri, A. I. U. Ivyside, Altoona, Pa.
Anthony, Brother Jerome, St. Marys College, Winona, Minn.
Asendorf, Robert H., 3544 N. Sydenham St., Philadelphia 40, Pa.
Attaya, William Lawrence (*J*), 14 Grew Ave., Roslindale 31, Mass.
Berger, Robert Lewis (*J*), Park College, Dept. of Physics, Parkville, Mo.
Barnothy, Jenö Michael, Barat College, 700 Westleigh Rd., Lake Forest, Ill.
Bottger, J. Edward (*J*), 17½ South First St., Evansville, Wisc.
Brouillette, Joseph Walter Jr., 16 S. St., Columbia, Mo.
Brown, Harry Allen (*J*), 215 S. Randall St., Madison, Wisc.
Buchta, John Charles (*J*), 807 West Nevada, Urbana, Ill.
Burrus, Charles Andrew Jr. (*J*), Box 1355, Emory University, Ga.
Carcelli, John Peter (*J*), 309 Mezes Hall, N. Brother Island, Bronx 53, N. Y.
Carruthers, Charles Woodside, R.F.D. 1, Brunswick, Maine
Castle, John G. Jr., 34 Hillcrest Dr., Clarence, N. Y.
Chabai, Albert John (*J*), 1120 South Sixth St., Bozeman, Mont.
Countee, Thomas Hilaire, 416 Luray Place, N. W., Washington, D. C.
Cowin, John Walter, 133 W. Pewabic St., Ironwood, Mich.
Curtis, Harold Ormand, Dept. of Physics, Hamilton College, Clinton, N. Y.
Dickinson, J. D. (*J*), 3416 E. Kellis St., Fort Worth, Texas.
Elliott, Paul P., Route 5, Lubbock, Texas
Elliott, Stuart Bruce (*J*), 311 Bonita Ave., Piedmont, Calif.
Engle, Leon Byron (*J*), 1113 Michigan Ave., Alamogordo, N. M.
Epstein, Saul Theodore, 403 W. 115 St., New York, N. Y.
Farrand, Weston B., 822 S. Willson St., Bozeman, Mont.
Forgy, Roline Alexander (*J*), 709 West 22nd St., Austin, Texas
Friedman, Harold Sidney (*J*), 10066 Cedarlawn, Detroit 4, Mich.
George, Wayland Dale (*J*), Rt. 1, Box 259, Strathmore, Calif.

Gilboy, Lawrence Daniel, 237 Curry Place, Youngstown, Ohio
Goldowski, Nathalie Michel, Black Mountain College, Black Mountain, N. C.
Goosey, Malcolm Hayes Jr. (*J*), 1003 W. Cleveland St., Bozeman, Mont.
Grosso, John S. (*J*), 514 Van Nest Ave., New York, N. Y.
Griffing, David F., 210 W. Spring St., Oxford, Ohio
Hagan, William Kelly (*J*), 116 Hamilton Park, Lexington, Ky.
Heaton, LeRoy, 1407 Walnut St., Columbia, Mo.
Hudson, Orlo K., 1524 Manor Rd., Austin, Texas
Innis, David Thomas (*J*), 231 N. Pearl St., Granville, Ohio
Johnson, Herman J., 501 Edgewood Rd., Flowerfield, Lombard, Ill.
Jones, Ernest Addison, Gilman Ave., Nashville, Tenn.
Jones, Robert Edward, R.D. 1, State College, Pa.
Joseph, Sister Mary, I.H.M., Library, Marygrove College, 8425 McNichols Rd., Detroit 21, Mich.
Kenney, Vincent Paul, Box 300, Fordham University, Bronx 58, N. Y.
Kleiner, Charles Theodore (*J*), 3402 Morehead St., El Paso, Texas
Knudsen, William Claire (*J*), 1014 Colby St., Madison, Wisc.
Kovach, Rev. Edward Michael, 352 Riverside Dr., New York 25, N. Y.
Lawrence, David Hughes, 210 W. Mt. Pleasant Ave., Philadelphia 19, Pa.
Linder, Solomon Leon (*J*), Dept. of Physics, Washington University, St. Louis, Mo.
Long, Robert Warren, 7318 LeHavre Rd., Houston 21, Texas
Lowe, Benjamin Sparkman, Box 265, Virginia State College, Petersburg, Va.
McGehee, Robert G. (*J*), Box 1355, Emory University, Ga.
McKay, Vern A. (*J*), Box 6225 North Texas Station, Denton, Texas
Macy, Spencer, 223 St. Charles, Rapid City, S. D.
Major, Schwab S. Jr., 1711 N. Vassar St., Wichita 14, Kan.
Malmberg, Paul R., 412 Whitney Ave., Pittsburgh 21, Pa.
Marinaccio, Lawrence Francisco (*J*), 228 Park Ave., Ellwood City, Pa.
Martin, Robert Keith, 422 N. Blanche St., Madison, S. D.
Miller, Darrell W., R. 1, Gaston, Ind.
Mount-Campbell, Paul, 1502 N. Missouri St., Roswell, N. M.
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## Colloquium of College Physicists

The Colloquium of College Physicists will have its next sessions on June 13–16 at the University of Iowa. There will be research lectures on low velocity electron beams, physical behavior of high polymers, semiconduction, acoustics and mass spectrometry. An afternoon will be occupied by special studies of undergraduate laboratory instruction, both advanced and elementary. There will be an exhibit of new experimental and nonexperimental teaching devices for which prizes are awarded. Professor George Gamow will give a series of four lectures, two on each of the subjects "Origin and Evolution of the Universe" and "Physics of Living Matter." Inquiries should be directed to Dr. G. W. Stewart, Department of Physics, University of Iowa, Iowa City, Iowa.